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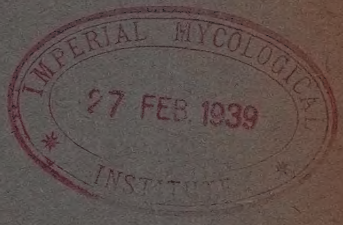

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MATHÉMATIQUES

SUR LES SYSTÈMES DE COURBES REMPLISSANT UN ESPACE MÉTRIQUE

Par W. NIEMYTZKI

(Présenté par S. N. Bernstein, de l'Académie, le 9.IX.1938)

1. Soit $\mathfrak{M}$ un espace métrique complet. Appelons «courbe» dans cet espace tout ensemble de points qui est une image biunivoque et continue d'une droite infinie. Cette droite sera nommée l'axe t et la courbe sera désignée par $f(p, t)$, où p est un certain point de l'espace situé sur la courbe considérée. On peut obtenir des courbes remplissant l'espace en résolvant des équations différentielles ordinaires, en résolvant le problème de Cauchy pour les équations aux dérivées partielles et, en général, en résolvant des équations fonctionnelles de la forme

$$\varphi = V(p, \varphi),$$

où l'on suppose que p étant donné, on connaît les conditions initiales.

Soit $\mathfrak{M}$ un espace métrique complet; on suppose qu'il passe par certains points de cet espace des courbes au sens précédemment défini; chaque point de l'espace $\mathfrak{M}$, par lequel il ne passe aucune courbe ou bien il passe plusieurs courbes, sera nommé point singulier et nous convenons de dire que les courbes qui passent par ces points dégénérent en points *. Remarquons ensuite que chaque arc ouvert d'une «courbe» au sens précédemment défini peut aussi être considéré comme une «courbe».

D'ailleurs, on considère comme points singuliers tous les points dont chaque voisinage contient des demi-courbes de la famille considérée. Remarquons qu'il suit de cette dernière condition que les points extrêmes des courbes sont encore des points singuliers. Une famille de courbes, dont certains points seront considérés comme singuliers d'après les conditions précédemment indiquées, est nommée famille S.

Nous dirons qu'une famille de courbes est «géométriquement continue» si, quel que soit $\varepsilon > 0$ et quel que soit l'arc $\widehat{ab}$ et le point q sur cet arc, on peut trouver un nombre δ tel que, si $\rho(c, q) \leqslant \delta$, il existe sur la courbe passant par le point c un arc $\widehat{d'cd}$ dont l'écart (1) de l'arc $\widehat{ab}$ est inférieur ou égal à ε.

---

* Si plusieurs courbes passent par un point, nous convenons de considérer chacune de ces courbes comme divisée en trois parties: deux demi-courbes et un point, ce point étant considéré comme une courbe dégénérée.

2. Soit $f(p, t)$ une certaine courbe; l'ensemble $\hat{\varepsilon}_f$ sera nommé ensemble de ses points-limites dynamiques si, quel que soit le point $q \subset \hat{\varepsilon}_f$, il existe une suite de valeurs du paramètre t, soit $t_1, t_2, \ldots t_n, \ldots$, qui tend vers $+\infty$ ou bien vers $-\infty$ et telle que

$$\lim_{n\to\infty} \rho(f(p, t_n), q) = 0.$$

De manière analogue à celle de la théorie de Birkhoff on distingue parmi les points de $\hat{\varepsilon}_f$ les points α-limites et ω-limites dont les ensembles seront désignés respectivement par $\hat{\varepsilon}_{f\alpha}$ et $\hat{\varepsilon}_{f\omega}$.

Théorème 1. *S'il existe un point de la courbe $f(p, t)$ qui appartient à $\hat{\varepsilon}_{f\omega}$ ($\hat{\varepsilon}_{f\alpha}$), tous les autres points de cette courbe appartiennent aussi à $\hat{\varepsilon}_{f\omega}$ ($\hat{\varepsilon}_{f\alpha}$).*

Soit q un point de $f(p, t)$ qui appartient à $\hat{\varepsilon}_{f\omega}$; il existe donc une suite $q_1 = f(p, t_1)$, $q_2 = f(p, t_2)$, $\ldots$ $q_n = f(p, t_n)$, $\ldots$ qui converge vers q et telle que $t_n \to +\infty$. Prenons sur la trajectoire $f(p, t)$ un point arbitraire r et considérons l'arc $\overset{\frown}{qr}$. Soit $\rho(q_1, qr) = \varepsilon_1$. Prenons un η_1 assez petit pour que dans le η_1—voisinage du point q—chaque point soit l'origine d'un arc $\overset{\frown}{q_{n_1} r_{n_1}}$ dont l'écart de $\overset{\frown}{qr}$ est inférieur à $\frac{\varepsilon_1}{2}$ et dont l'extrêmité se trouve à distance inférieure à $\frac{\varepsilon_1}{2}$ du point r. Ceci est toujours possible en vertu du lemme suivant.

Soit $\varepsilon > 0$; on peut prendre η assez petit pour que, si $\rho(x, y) \leqslant \eta \leqslant \frac{\varepsilon}{2}$, le diamètre de l'arc $\overset{\frown}{xy}$ soit $\leqslant \frac{\varepsilon}{2}$, c'est-à-dire $d(\overset{\frown}{xy}) \leqslant \frac{\varepsilon}{2}$. Considérons un $\frac{\eta}{4}$, voisinage de l'arc $\overset{\frown}{qr}$.

Alors quel que soit l'arc $\overset{\frown}{ab}$, situé à l'intérieur de cette sphère et tel que $\rho(a, q) \leqslant \frac{\eta}{4}$ et $\rho(b, q) \leqslant \frac{\eta}{4}$, son écart de l'arc $\overset{\frown}{qr}$ est $\leqslant \varepsilon$. Considérons la valeur du paramètre pour le point r_{n_1}. Soit $r_{n_1} = f(p_1, T_{n_1})$. Démontrons que $T_{n_1} > t_1$, où $q_1 = f(p_1, T_{n_1})$. En effet, si l'on avait $T_{n_1} > t_{n_1}$ [où $q_{n_1} = f(p, T_{n_1})$], comme $t_{n_1} > t_1$, on aurait $T_{n_1} > t_1$. Soit $T_{n_1} < t_{n_1}$; si d'ailleurs on avait $T_{n_1} < t_1$, l'arc $\overset{\frown}{q_{n_1} r_{n_1}}$ devrait contenir le point correspondant à la valeur du paramètre t_1, donc le point q_1, mais $\rho(q, \overset{\frown}{qr}) = \varepsilon_1 > \frac{\varepsilon_1}{2}$, et nous serions en contradiction avec la condition

$$\alpha(\overset{\frown}{qr}, \overset{\frown}{q_{n_1} r_{n_1}}) \leqslant \frac{\varepsilon_1}{2}.$$

Ceci étant établi, il est facile d'achever la démonstration.

La continuité géométrique nous permet ensuite de démontrer que les ensembles $\hat{\varepsilon}_{f\alpha}$ et $\hat{\varepsilon}_{f\omega}$ sont des ensembles invariants, c'est-à-dire des ensembles formés de courbes entières de la famille. Pour le voir, nous allons d'abord déduire une formule qui permet de définir ces ensembles indépendemment du choix du paramètre t sur les courbes.

Théorème 2. *Si la courbe $f(p, t)$ n'est ni un point singulier ni une trajectoire fermée, l'ensemble des points-limites dynamiques peut être défini par la formule:*

$$\varepsilon f = \hat{\varepsilon}_f = \overline{f - f}.$$

La démonstration de l'inégalité $\hat{\varepsilon}_f \subset \varepsilon_f$ est triviale. Démontrons l'inégalité inverse. Soit $p_0 \subset \hat{\varepsilon}_f$; alors ou bien $p_0 \subset \overline{f} - f$, et il est alors évident que $p_0 \subset \hat{\varepsilon}_f$, ou bien $p_0 \subset \hat{\varepsilon}_f$ et $p_0 \subset f$, mais alors $f(p, t) \subset \hat{\varepsilon}_f$ en vertu du théorème 1.

Comme p_0 est un point ordinaire, il existe une sphère $S = S(p_0, d)$ telle que chaque arc de la famille des courbes appartenant à $\overline{S}$ a une longueur finie par rapport au temps, c'est-à-dire présente l'image d'un segment de l'axe t. Désignons par A l'intersection $f\overline{S}$. Alors $A = f\overline{S}$ se décompose en une infinité dénombrable d'arcs L_j différents ($j = 1, 2, 3, \ldots$) de la courbe f; chacun de ces arcs est évidemment un ensemble fermé. Considérons l'ensemble $L_j - (\overline{S} - S)$; il est formé d'un nombre fini ou d'une infinité dénombrable d'arcs ouverts L_j^k, c'est-à-dire

$$\sum_i \sum_k \overline{L_j^k} \subset f\overline{S} = A.$$

Nous démontrons ensuite que chaque arc $\overline{L_j^k}$ est partout non dense sur l'ensemble $\overline{A}$, c'est-à-dire

$$\overline{A} = \overline{\overline{A} - L_j^k}.$$

Comme $\sum_i \sum_j \overline{L_j^k}$ est la somme d'une infinité dénombrable au plus d'ensembles fermés, l'ensemble $\sum_i \sum_j \overline{L_j^k}$ est un ensemble de première catégorie sur $\overline{A}$.

D'autre part, $\overline{A}$ est un ensemble fermé situé dans un espace métrique complet, donc il ne peut pas être de première catégorie sur lui-même et il en résulte

$$\overline{\overline{A} - \sum_j \sum_k \overline{L_j^k}} = \overline{A}.$$

Il est maintenant facile à montrer que

$$p_0 \subset \overline{\overline{f} - f} = \varepsilon_f.$$

Remarquons que si l'on désigne par f_+ et f_- respectivement les demi-trajectoires positive et négative, on a

$$\hat{\varepsilon}_{f\alpha} = \overline{\overline{f}_+ - f_+}$$

$$\hat{\varepsilon}_{f\omega} = \overline{\overline{f}_- - f_-}.$$

En s'appuyant sur ces théorèmes, on établit facilement les résultats suivants:

Corollaire 1. L'ensemble $\hat{\varepsilon}_f$ est un ensemble invariant. En effet, soit $q \subset \hat{\varepsilon}_f$, alors $q \subset \overline{f}$ et q n'appartient pas à f, donc $q \subset \overline{f} - f$. Soit r un point arbitraire de $f(q, t)$. En vertu de la continuité géométrique nous voyons que $r \subset \overline{f}$, et comme r n'appartient pas à f, on a $r \subset \overline{f} - f$, donc $r \subset \hat{\varepsilon}_f$.

Corollaire 2. Les ensembles $\hat{\varepsilon}_{f\omega}$ et $\hat{\varepsilon}_{f\alpha}$ sont formés de courbes entières.

3. Définition 1. Une demi-courbe est nommée stable au sens de Lagrange si elle est entièrement contenue dans un certain ensemble compact.

Théorème 3. *Si* $\bar{f}=f$, *alors ou bien* f *est un point, ou bien une trajectoire fermée, ou bien chaque demi-trajectoire de* f *est non stable au sens de Lagrange.*

Définition 2. Nous dirons qu'une courbe $f(p, t)$ possède un col à l'infini[2] s'il existe un ensemble C compact et fermé tel que, C_a étant un autre ensemble compact et d'ailleurs arbitraire, il existe dans C deux points $\bar{p}$ et $\bar{q}$ situés sur la trajectoire $f(p, t)$ et tels que l'arc $\bar{p}\bar{q}$ possède des points hors de C_a.

Théorème 4. *Pour qu'une trajectoire non stable au sens de Lagrange soit un ensemble fermé il est nécessaire et suffisant qu'elle ne possède pas de col à l'infini.*

4. Définition 1. Une courbe $f(p, t)$ est nommée stable au sens de Poisson, si ses points sont des points-limites dynamiques pour elle-même, c'est-à-dire si

$$f \subset \overline{\overline{f}-f}.$$

Définition 2. Un ensemble A de courbes est appelé minime, si pour chacune des demi-courbes f_+ et f_- qui le composent on a l'égalité

$$\bar{f}_+ = \bar{f}_- = A.$$

Le rôle des ensembles minimes et leur relation avec les courbes stables au sens de Poisson sont caractérisés par les deux théorèmes suivants:

Théorème 5. *L'ensemble des points-limites dynamiques d'une courbe stable au sens de Poisson au moins dans une direction contient un ensemble minime.*

Théorème 6. *Si un ensemble minime est compact, toutes les courbes dont il est formé sont stables au sens de Poisson.*

Institut de Mathématiques.
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MATHÉMATIQUES

SUR LES FAMILLES DE COURBES DU TYPE DE BENDIXON

Par W. NIEMYTZKI

(Présenté par S. N. Bernstein, de l'Académie, le 9. IX.1938)

1. Dans l'étude de la disposition dans le plan des courbes présentant les solutions des équations différentielles ou fonctionnelles les faits essentiels sont résumés dans les théorèmes de Poincaré-Bendixon exposés dans leurs mémoires classiques. Le but de la présente Note est de donner une caractéristique locale pour les familles de courbes qui vérifient encore les théorèmes de Bendixon.

2. Introduisons une définition. Nous dirons qu'une famille de courbes est une famille du type de Bendixon, si les conditions suivantes sont vérifiées:

I. Il passe par chaque point du plan une et une seule courbe. Certaines de ces courbes peuvent dégénérer en points, ces points seront nommés singuliers.

II. Soit G un domaine qui contient un point singulier, alors chaque demi-courbe entièrement située dans ce domaine est

1) ou bien une courbe fermée entourant le point singulier,
2) ou bien elle a ce point singulier pour point-limite unique,
3) ou bien elle est une spirale qui s'enroule autour d'une courbe fermée entourant le point singulier,
4) ou bien elle est une spirale qui s'enroule autour d'une courbe fermée passant par le point singulier.

Remarquons qu'il suit de ces conditions que

a) si un point limite se trouve situé sur une courbe, elle est nécessairement fermée;
b) il existe au moins un point singulier à l'intérieur de chaque courbe fermée.

3. En suivant M. Kerekjarto (1), nous dirons q'un point $\overline{a}$ du plan est un point régulier d'une famille, s'il existe un voisinage (un domaine) du point $\overline{a}$ tel que tous les segments des courbes remplissant ce domaine peuvent être transformés simultanément en segments de droites au moyen d'une certaine transformation biunivoque réciproque et continue. Un tel voisinage sera nommé dans la suite «voisinage assez petit».

Théorème I. *Si une famille de courbes ne possède pas de points singuliers et si chacun de ses points est régulier, alors les courbes de cette famille peuvent être des trajectoires d'un certain groupe continu à un paramètre, donc cette famille est du type de Bendixon.*

Considérons un arc $\widehat{pq}$ arbitraire et entourons chacun de ses points par un voisinage «assez petit». Ces voisinages sont supposés sphériques et l'on en choisit un nombre fini de manière à couvrir l'arc entier et de sorte que le centre de chaque cercle suivant se trouve sur la circonférence du cercle précédent. Établissons maintenant sur l'arc $\widehat{pq}$ une certaine direction et prolongeons-la sur toute la courbe $f(p, t)$ de la famille. Sur tous les arcs du premier voisinage où la direction n'a pas encore été établie [elle aurait pu être établie si les arcs étaient des morceaux de la courbe $f(p, t)$], on établit une direction en accord avec celle de $\widehat{pq}$*. Après avoir établi la direction sur un certain segment, on la prolonge sur toute la courbe dont le segment est une partie. On passe ensuite du premier voisinage au second et ainsi de suite jusqu'à ce qu'on arrive au voisinage du point q. Montrons maintenant que tous les arcs d'un petit voisinage de $\widehat{pq}$ seront orientés en accord.

En effet, considérons un point arbitraire r sur l'arc $\widehat{pq}$; comme r est régulier, on peut trouver un arc $\widehat{\alpha\beta}$ tel que chaque arc d'un voisinage assez petit du point r coupe l'arc $\widehat{\alpha\beta}$. Si les directions sur les arcs n'étaient pas en accord, une certaine courbe de notre famille couperait l'arc $\widehat{\alpha\beta}$ sans contact dans deux directions opposées, et dans ce cas, en vertu d'un principe énoncé déjà par Poincaré et démontré par exemple dans le livre de Kerekjarto (2), il existerait un point singulier, ce qui contredit aux conditions du théorème.

Couvrons maintenant tout le plan par une infinité dénombrable de voisinages assez petits et prolongeons l'orientation sur toutes les courbes. Or, si une famille de courbes est formée de points réguliers seulement et si elle admet une orientation, elle peut, ainsi qu'il a été démontré par Whitney (4), être considérée comme une famille de trajectoires d'un certain groupe continu à un paramètre.

4. Supposons maintenant que la famille possède des points singuliers, mais que tous les autres points sont réguliers. Introduisons la définition suivante: une famille de courbes sera nommée «géométriquement continue», si, quelque petit que soit $\varepsilon > 0$ et quel que soit l'arc $\widehat{ab}$ et le point q sur cet arc, on peut trouver un δ tel que, si $\rho(c, q) \leqslant \delta$, il existe sur la courbe passant par le point c un arc $\widehat{d'cd}$ dont l'écart de l'arc $\widehat{ab}$ est inférieur à ε.

Il est facile à voir que chaque famille dont tous les points, excepté les points singuliers, sont réguliers est géométriquement continue. Donc, si la condition d'unicité est remplie, on a le théorème:

L'ensemble des points ω-limites (α-limites) de chaque courbe est formé de courbes entières de la famille (3).

Théorème II. *Chaque famille vérifiant la condition d'unicité et dont chaque point, excepté les points singuliers, est régulier est une famille du type de Bendixon.*

Considérons une courbe qui entre dans le domaine ou bien qui commence dans le domaine et n'en sort pas; une telle courbe possède des points-limites dynamiques. Deux cas sont possibles: ou bien parmi ces

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* Quand nous transformons le voisinage «assez petit» de telle manière que tous les segments des courbes passant par ce voisinage se transforment en segments situés sur des droites parallèles, le choix de la direction sur l'arc détermine le choix de la direction sur tous les autres arcs de courbes.

points il y a des points singuliers, ou bien il n'y en a pas. Le premier cas est considéré dans le livre de Kerekjarto; en appliquant le principe de Poincaré déjà cité, il a démontré que la courbe est ou bien fermée, ou bien une spirale qui s'approche indéfiniment d'une courbe fermée. Supposons que parmi les points ω-limites (α-limites) de la courbe il y a un point singulier et supposons d'ailleurs qu'il y a encore d'autres points ω-limites, par exemple, le point q. Supposons, par impossible, qu'il existe parmi les points α-limites ou bien ω-limites de $f(q,t)$ un point S qui n'est pas singulier. En vertu du théorème précédemment cité, toute la courbe $f(s,t)$ est formée de points ω-limites pour la courbe $f(q,t)$. Menons (ce qui est possible en vertu de la régularité) par le point S un arc $\widehat{\alpha\beta}$ sans contact; alors on démontre comme ordinairement (Bendixon) que cet arc sans contact est coupé par une courbe L une infinité de fois et d'ailleurs dans deux directions opposées, mais dans ce cas, en vertu du principe de Poincaré, il existe une infinité de points singuliers à l'intérieur du domaine, ce qui contredit à l'hypothèse faite.

Il est facile à donner un exemple d'une famille du type de Bendixon dont les courbes ne peuvent pas être trajectoires d'un groupe continu à un paramètre. Telle est, par exemple, la famille des ovales concentriques qui s'approchent indéfiniment d'un segment.

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PHYSICS

ISOTOPE SHIFT OF SPECTRAL LINES AND THE INTERACTION OF NEUTRONS WITH ELECTRONS

By IG. TAMM, Corresponding Member of the Academy

The dependence of the energy levels of electrons on the mass of the nucleus (due to the motion of the nucleus and of the electrons relative to their common centre of inertia) accounts for the isotope shift in the spectra of the lightest elements only (such as H and Li), whereas the observed shift of some spectral lines of Tl, Hg and Pb is almost 10^3 times larger than this «mass effect» and of an opposite sign. Even in the case of such a light element as Ne the observed shift of some spectral terms (1) is decidedly larger than that expected on account of both the «normal» and the so-called «specific» effect (2) of the motion of the nucleus.

If one compares the isotope displacements of different terms of an element, one finds that the amount of the displacement increases with the number of penetrating electrons and the degree of penetration of the electrons in the neighbourhood of the nucleus (2). This indicates that the larger isotope shifts are due to some difference in the field of various isotopes in the neighbourhood of the nucleus.

It was pointed out by Bartlett in 1931 that since the heavier isotopes have greater radii, the electric field within their nuclei must be weaker and they must bind a penetrating electron more loosely than the lighter isotopes. Calculations carried out by Breit showed that in the case of heavy elements (Hg, Tl and Pb) this «radius effect» is of the right order of magnitude. However, in case of lighter elements the sign of the observed shift is opposite to the sign of the «radius shift». The shift is described as positive, if the terms of the lighter isotopes lie lower than those of the heavier isotopes, and as negative in the opposite case. Now the «radius shift» is positive, whereas the observed shift (or rather the difference between the observed shift and the shift expected on account of the mass effect) is negative for the lighter elements up to about $Z=35$, vanishes in the middle of the periodic table and becomes positive only for relatively large Z.

Thus, in addition to the difference caused by the variation of the radius there must exist some other difference in the fields of various isotopes in the neighbourhood of the nucleus. It seems probable that this difference is due to a short range interaction of electrons with the additional neutrons in the nuclei. Since this interaction has to account for the negative shifts in the spectra of the lighter elements, it must be equivalent to an a t t r a c t i o n between neutrons and electrons.

The existence of a weak neutron-electron interaction was first suggested by Condon (3). Using the data on the ionization produced by neutrons in

gases and on the scattering of slow neutrons, he determined an upper limit for the intensity of this interaction. Further he showed that a repulsion between neutrons and electrons, somewhat weaker than this upper limit, may explain even the largest isotope shifts in the spectra of Tl, Hg and Pb. However, Condon completely neglected the radius effect and did not try to explain the general dependence of the isotope shift on the atomic number and the change of the sign of this shift in the middle of the periodic table.

Let us assume with Condon a δ-type interaction between neutrons and electrons and let us suppose that this interaction may by described by an operator of the form

$$\Omega = -fO\Psi^*\Psi\psi^*\psi, \qquad (1)$$

where Ψ and ψ are the wave functions of the neutron and the electron, O denotes a relativistically invariant combination of these functions and f is a universal constant. Using for the neutron the non-relativistic approximation and averaging over the directions of its spin, one obtains

$$\Omega = -f|\Psi|^2[a\psi^*\psi + (1-a)\psi^*\beta\psi],$$

where β is one of the Dirac matrices and a—a constant depending on the choice of the operator O in (1). Since the results in which we are interested do not materially depend on the value of a, we will set in the following $a=1$:

$$\Omega = -f|\Psi|^2\psi^*\psi. \qquad (2)$$

Let us further assume that the short range interaction (1) and (2) exists also between protons and electrons*.

Suppose that the nuclear particles are uniformly distributed within a sphere of the radius $R=A^{1/3}R_0$, where $R_0 = \frac{r_0}{2} = \frac{e^2}{2mc^2}$ and A is the mass number of the nucleus. Then the isotope displacement caused by the addition to the nucleus of ΔA neutrons, due both to the radius effect and to the interaction (1), can be calculated with the help of a formula deduced by Rosenthal and Breit (4). The change of energy of an S-state of an electron is equal to

$$\Delta E = \psi^2(0)\frac{2(1+s)}{\Gamma^2(1+2s)}\left(\frac{2ZA^{\frac{1}{3}}R_0}{a_H}\right)^{2s-2}\left\{\frac{4\pi Ze^2R_0^2}{A^{\frac{1}{3}}(1+2s)(3+2s)} - f\right\}\Delta A. \qquad (3)$$

Here $\psi(0)$ is the value of the Schrödinger electron wave function at $r=0$, $s^2 = 1-\left(\frac{Ze^2}{hc}\right)^2$ and a_H is the hydrogen radius $\frac{h^2}{me^2}$.

The first term in the brackets accounts for the radius effect and increases with Z approximately as $Z^{2/3}$, while the second term, which accounts for the interaction (2), does not depend on Z. Therefore it is possible to choose the value of the constant f so as to account both for the negative shifts at the beginning of the periodic table and for the positive ones at the end of it.

The greatest uncertainty in (3) arises from the factor $\psi^2(0)$. In case of complex atoms only the order of magnitude of this factor can be determined theoretically, since the value of $\psi^2(0)$ is very sensitive to the perturbations of the wave function caused by the interaction of the electrons. It is therefore very fortunate that the value of the universal constant f may be estimated by a kind of a «zero method» independently from the determination

* If that is not the case, the second term in the brackets in the equation (3) has to be multiplied by $\left(1+\frac{3Z}{A(1+2s)}\right)$.

of $\psi^2(0)$. In fact, the vanishing of the shift in the middle of the periodic table corresponds to the vanishing of the expression in the brackets in (3) which is independent from $\psi^2(0)$.

To estimate roughly the dependence of the absolute value of the shift on Z, we have used the approximate expression:

$$\psi^2(0) = \frac{Z(1+z)^2}{\pi n^{*3} a_H^3}, \qquad (4)$$

where z is the degree of ionization of the atom and n^*—the effective total quantum number. The quantum number n^* in (4) was arbitrarily put equal to 1 and ΔA in (3) equal to 2. The Curves *1*, *2*, *3* on the Fig. 1 represent the results obtained for $f=0$, $f=2.1\ mc^2 r_0^3$ and $f=2.3\ mc^2 r_0^3$ respectively $\left(r_0 = \frac{e^2}{mc^2}\right)$. The circles on the figure represent the largest observed values of the shift in the spectra of all the elements* (beginning with Ne) for which data were available ([5]). We have substracted from the observed maximum shift the trivial shift due to the normal mass effect and divided the rest by $(1+z)^2$ (cf. equation 4). Only isotopes possessing identical angular nuclear momenta were included.

Fig. 1 shows that the radius effect alone (Curve *1*) cannot account for the isotope shifts of different elements. On the other hand, in vue of the roughness of our approximations [the very rough estimate (4) of the value of $\psi^2(0)$, made even more rough by arbitrarily putting n^* equal to 1; the neglect of the very important perturbation effects, etc.] the general agreement of our Curves *2* and *3* with the dependence of the observed shifts on Z seems to be satisfactory** and indicates that the true value of the universal constant f must be equal to about

$$f = 2.3 mc^2 r_0^3 = 4 \cdot 10^{-44}\ \text{erg/cm}^3. \qquad (5)$$

This value is 13 times smaller than the allowable upper limit of f determined by Condon ([3]) from the data on the scattering of slow neutrons.

In deducing the equation (2) we have averaged the interaction (1) over all possible directions of the spin of the neutron. Most probably the interaction (1) depends on the relative directions of the spins of the neutron (or of the proton) and of the electron, so that the hyperfine structure of spectral lines is influenced by this interaction. If the spin-dependent part of the interaction is of the same order of magnitude as the spin-independent one, the corresponding hyperfine separation of the S-terms must be about 20 g times smaller, than the separation caused by the magnetic moment of the nucleus (g denotes the Landé factor).

However, the spin interaction cannot displace the centre of gravity of the spectral line; the present theory can thus not account for the fact that the components of spectral lines corresponding to odd isotopes are usually nearer to the components of the next lighter even isotope than to those of the next heavier one***.

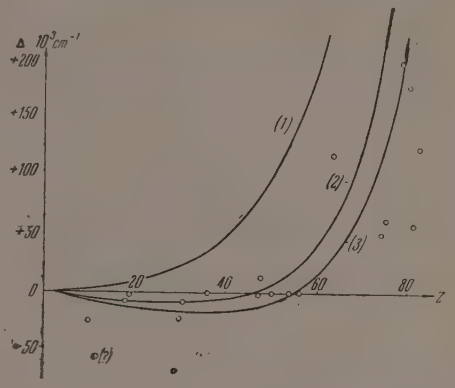
---

* We have left out only Sm since Schüler and Schmidt (Z. Physik, **92**, 148, 1934), who have measured the shift in the spectrum of Sm, do not state explicitly whether it was positive or negative. Fig. 1 of their paper seems to indicate that the shift is negative, whereas the reference in the text to the radius effect and the relatively large positive shift of the neighbouring element Eu suggest that the shift is positive. Anyhow the shift of Sm is anomalous, Sm being the only striking exception to the rule that the components of spectral lines due to e v e n isotopes are equally spaced. This peculiarity is probably connected with the α-radioactivity of Sm.

** It must be remembered that the experimental data themselves are not quite reliable.

*** The exceptionally large anomaly of the position of the component of the odd isotope Mg_{25} relative to those of the even ones Mg_{24} and Mg_{26}, found by Bacher and Sawyer, is disproved by the new measurements of K. Meissner (Phys. Rev., **53**, 931, 1936).

The interaction (1) implies the possibility for a neutron or a proton to create a positron-electron pair; to prove this one has only to consider the electron wave function ψ as quantized. Therefore an excited nucleus can loose its energy not only by γ-radiation, but also by the direct emission of a pair. However, up to extremely high excitation energies (of the order of 10^8 eV) this process is by far less probable than the emission of a γ-ray by the nucleus, so that the number of «directly emitted» pairs must be quite negligible in comparison with the number of pairs created by the internal conversion of γ-rays.



The possibility for a heavy particle to create a pair implies the existence of an interaction between the heavy particles due to the exchange of pairs between them. It was first suggested by Teller [6] and by Wentzel [7] that such an exchange of pairs may account for the nuclear forces. However, with the value (5) of f the energy of interaction $V(r)$ of two heavy particles, separated by a distance r, turns out to be far too small:

$$V(r) = -100 \left(\frac{r_0}{r}\right)^5 eV. \qquad (6)$$

Recently I have suggested [8] that an interaction of the type (1) may, perhaps, exist between neutrons and protons, on one hand, and between both the ordinary and the heavy electrons, on the other. Trying to explain with the help of this assumption some properties of the penetrating cosmic rays as well as the nuclear forces, I arrived at a value of the universal constant f about 10 times larger than the value (5). This discrepancy may be due to the fact that, contrary to the assumption made in the calculations just mentioned, heavy electrons do interact with neutrons and protons stronger than ordinary ones. At any rate applying the present form of the quantum theory to problems involving small distances of the order of r_0 and high energy transfers, one cannot expect to obtain more than rough estimates.

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PHYSICS

PRIMARY IONIZATION WITH FAST ELECTRONS IN NITROGEN*

By M. D. BORISSOV, V. P. BRAILOVSKI and A. I. LEIPUNSKI

(Communicated by S. I. Wawilow, Member of the Academy, 21. IX. 1938)

In passing through matter fast electrons spend a part of their energy on ionizing the atoms. The number of acts of ionization per unit path performed by a primary electron is called magnitude of primary ionization.

Examining the interaction of a fast electron with hydrogen atoms with the aid of quantum mechanics and taking into account the relativistic properties of the electron, Bethe (1) and later Möller (2) and Williams (3) found for the primary ionization in atomic hydrogen the following expression

$$I=\frac{2\pi ne^4}{mv^2J}0.285\left[\log\frac{42\,mv^2}{J(1-\beta^2)}-\beta^2\right], \qquad (1)$$

where $\beta=v/c$.

This formula predicts that the primary ionization will diminish with the electron energy increasing to 1.5×10^6 eV, after which it will slowly increase and reach for the energy of the order of 10^9 eV a value 70% greater than the minimum value at 1.5×10^6 eV.

It may be safely affirmed that the general form of the dependence of the ionization on the electron energy will remain for other substances the same as in formula (1), although from the existing theory one can predict for the absolute value of the ionization the order of magnitude only.

Very little experimental work has been done on measuring the primary ionization and the results are not definite enough.

The primary ionization is usually measured by counting the number of drops on the photographs of the tracks of electrons in a Wilson cloud chamber. Thus, the number obtained must be somewhat larger than the real primary ionization, since it includes, beside the drops produced by the primary electrons, some secondary ions produced by the electrons detached from atoms by the primary particles. Under different experimental conditions the number of counted secondary ions is different, and this seems to account for the considerable discrepancies among the absolute values of the primary ionization given by different authors.

Wilson was the first to determine the order of magnitude of the primary ionization by fast electrons by counting the drops on 9 tracks obtained in the cloud chamber. The speed of the electrons was not measured, but it was

* This paper was read before the meeting of the Ukrainian Physico-Technical Institute at 19. I. 1937, but the publication of it was accidentally delayed.

estimated to be greater than 2×10^9 cm/sec. Wilson estimated the primary ionization in air to 18—22 ions/cm.

Terroux and Williams (4), placing the cloud chamber in a magnetic field, found the dependence of the primary ionization on the energy of the electrons on 72 tracks in oxygen and the same on 108 tracks in hydrogen. Both for oxygen and hydrogen they found that the primary ionization first dropped off rapidly with increasing electron energy and then tended to a certain minimum value. For the electron energy 1.6×10^6 eV (maximum electron energy in these experiments) the authors found the primary ionization in oxygen to be 22 ions/cm and in hydrogen for the 1.12×10^6 eV (maximum energy) 5.1 ions/cm.

Skramstad and Loughridge (5), working with electrons up to 2×10^6 eV, found the dependence of the primary ionization in nitrogen on the energy of the electrons on 100 tracks. For the electron energy of 1.5×10^6 eV the authors found in nitrogen 19.3 ions/cm. For energies greater than 1.5×10^6 eV no increase of the primary ionization was observed by the authors. Finally, for the region of ultra-fast electrons one may cite the works of Anderson (6), who found on diffusion tracks 31 pairs of ions/cm, and the work of Kunze (7), who obtained on sharp tracks 20 pairs of ions/cm.

Thus, one can hardly derive from the experiments mentioned any definite idea as to the behaviour of the primary ionization as a function of the electron energy.

In this work are reported the results of the measurements of the primary ionization obtained from a pretty large number of photographs of tracks of fast electrons in nitrogen.

An automatic cloud chamber 15 cm in diameter was filled with nitrogen and water and alcohol vapours. On the inside wall of the cloud chamber was fastened a platinum foil which was radiated by gamma-rays from an active thorium deposit. The small ampulla containing the source of 0.5 millicuries was surrounded on all sides with a 10-cm layer of lead. On the side of the cloud chamber there was a channel in the lead, in front of which the ampulla was placed at the time of photographing.

Two photographs were taken at right angles by means of two FED cameras. The cloud chamber was illuminated by 1000 W, 120 V lamps which were connected to 220 V before the expansion. The tracks were reproduced by projection on a mobile screen by means of the optical system which had served for photographing. The radii of curvature of the tracks were measured by superposing on them arcs of various radii drawn on a sheet of celluloid.

From all the tracks we selected such as: 1) were at least 5 cm in length, 2) had uniform curvature throughout the path and 3) had a distribution of drops along the path of the electron permitting to count. The result was reduced to normal conditions and corrected for the coalescence of drops, according to Kunze's formula (7):

$$N_0 = N\frac{N\varepsilon}{1-e^{-N\varepsilon}}.$$

N_0 is the real number of drops per unit path, N—the observed number of drops per unit path, ε—the diameter of the drop. The diameter of the drop on our photographs was about 0.4 mm. We have taken 1000 pairs of pictures from which we selected and studied 264 tracks.

In order to construct the curve we computed the mean value of I, i. e. the average number of drops per centimeter for each group of tracks whose energies lay in a certain definite interval $\Delta H\rho$. This mean value was attributed to the electron whose energy corresponded to the middle point of the interval $\Delta H\rho$. The result of this construction is shown in Fig. 1. The inter-

val $\Delta H\rho$ was chosen to 1000 gauss-cm. The figures over the points of the curve indicate the number of tracks in the respective group. The points denoted by crosses have been obtained at the same value of $\Delta H\rho$, but all the intervals are shifted by 500 gauss-cm to the right with respect to the first construction.

As seen from Fig. 1, our curve has a minimum corresponding to the electron energy of 1.5×10^6 eV. The fact that the points denoted by crosses and those denoted by circles give the same curve shows that the minimum of the curve cannot be explained by the arbitrary choice of the interval $\Delta H\rho$.

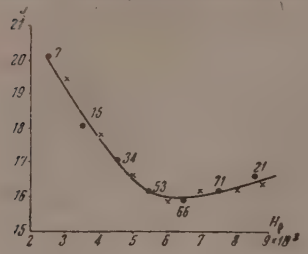

Fig. 1.

It must be noted that, owing to the presence of the hard intense line 2.62×10^6 eV in our source, the major part of the selected tracks fall on the most interesting portion of the curve.

In Fig. 2 are shown the curves obtained by different authors. These curves show that Terroux and Williams could not find the minimum since their source furnished electrons not above 1.5×10^6 eV. Skramstad and Loughridge had one point with the energy greater than 1.5×10^6 eV. But the reliability

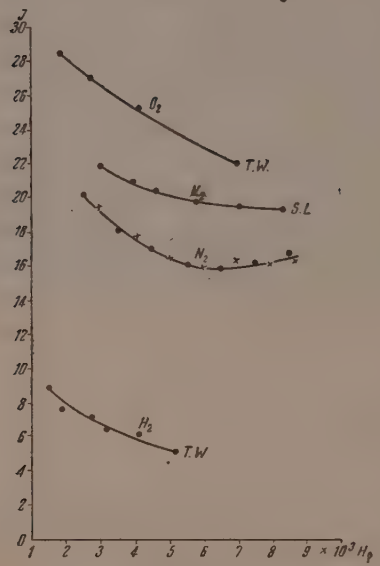

Fig. 2.

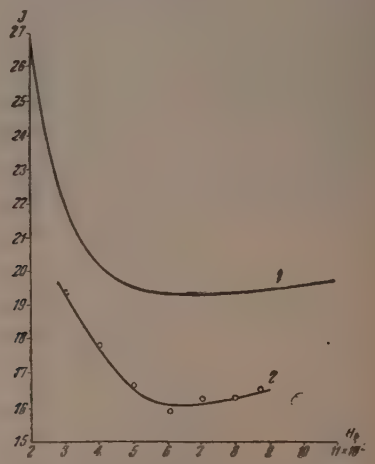

Fig. 3. *1*—theoretical curve, 2—experimental curve.

of this point is difficult to estimate, the authors giving no number of tracks corresponding to this point.

Fig. 2 gives an idea of the absolute values of the primary ionization obtained by different authors. The curve obtained by us gives the least value of I. Generally speaking, the numerical value of I can be largely affected by the following factors: 1) coalescence of drops and 2) secondary ions. While the first phenomenon can be approximately taken into account, this can hardly be done in the latter case. We endeavoured to minimize the influence of the secondary ions. To this purpose the platinum foil in the cloud chamber was irradiated at the end of the expansion when the

cloud chamber was on the point of forming drops. Owing to this, the processes of formation of ions and drops took place at practically the same time and the diffusion of ions could be neglected.

In the absence of the diffusion the distorting effect of the secondary ions must be very feeble, as the length of their path is less than the diameter of the drop.

Fig. 3 represents the curve obtained by us and the curve calculated from equation (1) on the assumption that $n=10\ n_0$, where n is the number of electrons in 1 cm^3, n_0 is the number of molecules in 1 cm^3 and $J=16.6$ eV.

Although one cannot expect a complete coincidence of the theoretical curve computed from equation (1) with the experimental curve obtained by us in nitrogen, yet Fig. 3 shows that these curves have a similar shape and that the increase of the primary ionization for electrons with energies above 1.5×10^6 eV is actually observed.

Thus, we are led to the conclusion that the quantum theory of the primary ionization predicts correctly the general character of the change of the primary ionization with the electron velocity.

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PHYSICS

HEAT TRANSFER OF GAS TO WALLS OF VESSEL BY FREE CONVECTION

By G. A. BARSKI and J. B. ZELDOVICH

(Communicated by N. N. Semenov, Member of the Academy, 13.IX.1938)

In course of February and March 1938 we carried out a series of experiments on the heat exchange between a gas contained in a vessel and the walls of this vessel.

The vessel used by us was a steel cylindrical bomb with an internal diameter of 175 mm, an internal length of 320 mm and a wall thickness of 7.5 mm, superimposed on the internal circumference.

After the vessel was filled with the gas under investigation (air, hydrogen), a part of this gas was allowed to escape through a quick-acting valve. In 0.01—0.02 sec. the valve was shut and the gas, now cooled through adiabatic expansion, was heated by heat exchange with the walls of the vessel, the process continuing till initial room temperature corresponding to the temperature of the walls was reestablished.

The course of the process under investigation could be observed by the curve of change in pressure with time. The pressure was recorded by light thrown off by a small mirror which was connected to a steel or glass membrane, mounted into the side surface of the vessel on a reel of photosensitive paper made to rotate at a certain fixed rate.

The principles of our experimental procedure were borrowed from the old work conducted by Witz (1), which may be said to have been somewhat imperfect from the point of view of the methods used.

Initial pressure varied from 8 atm. to 260 mm mercury column; while final pressure constituted from 57 to 89% of the initial one, which corresponded to a temperature drop of 58 to 14°. The experimental material was treated on the assumption that the coefficient of heat transfer remained constant throughout every individual experiment. Accordingly this material is described by interpolation formula

$$\lg Nu = 0.10 + 0.235 \lg Gr, \qquad (1)$$

found on the basis of the least squares method. In the interval investigated $\lg Gr$=from 9.5 to 5.2. This formula does not practically differ from the mnemonic:

$$\lg Nu = 0.25 \lg Gr; \quad Nu = \sqrt[4]{Gr}. \qquad (2)$$

In formulae (1) and (2) all material constants are referred to the final state, i. e. the temperature of the vessel walls.

Considered separately, the data of the experiments with hydrogen and air fail to disclose any systematic deviations.

The dependence $Nu=f(Gr)$, not investigated by us, with small values of Gr may be predicted with great probability by confronting formulae (1) and (2) with the terminal formula deduced theoretically(2):

$$Nu = Nu_0 + CGr^2. \qquad (3)$$

The value of Nu_0 can be easily calculated; it equals in our vessel 5.66. From the condition of the continuity of Nusselt's criterion as a function of Grashoff's criterion and its first derivative in the point where formulae (2) and (3) converge, it becomes possible both to find the very point of convergence and the magnitude of the coefficient in formula (3).

For cylindrical vessels with ratio of length of internal circumference to diameter 2 : 1 the ultimately recommended formulae will have the following appearance:

$$\left.\begin{aligned} Nu &= Nu_0 + 7^7 \cdot 8^{-8} Nu_0^{-7} \cdot Gr^2, & Gr &\leqslant 8^4 \cdot 7^{-4} Nu_0^4 \\ Nu &= 5.66 + 0.26 \cdot 10^{-6} Gr^2, & Gr &\leqslant 1750 \end{aligned}\right\} \qquad (4)$$

$$Nu = Gr^{\frac{1}{4}} \qquad Gr \geqslant 1750 \qquad (5)$$

Convection makes heat transfer double as compared with pure conductivity with $Gr=(2Nu_0)^4=1.7 \cdot 10^4$.

Thus, for instance, this doubling of heat transfer is obtained by cooling air in horizontal tubes with a difference in temperature between walls and air equalling 10°; walls being of room temperature with atmospheric pressure with a diameter of 8.5 mm and 100 mm mercury column with a diameter of 34 mm.

The above-mentioned figures permit of judging of the part of free convection under the conditions of laboratory experiments, for instance, in studying a heat explosion.

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PHYSICS

ABSOLUTE OUTPUT OF RADIATION CAUSED BY ELECTRONS MOVING WITHIN A MEDIUM WITH SUPER-LIGHT VELOCITY

By P. A. CHERENKOV

(Communicated by S. I. Wawilow, Member of the Academy, 28.IX.1938)

1. In the study of radiation arising in a substance due to the passage of fast electrons through it (1) the question of the absolute output of this radiation is of particular importance. According to Tamm and Frank's theory (2), energy W, emitted by an electron moving uniformly within a medium, is determined by the following expression:

$$W=\frac{e^2 l}{c^2}\int \omega d\omega \left(1-\frac{1}{\beta^2 n^2}\right), \qquad (1)$$

where l is length of electron path, n—refractive index of the medium, $\omega=2\pi\nu$—cyclic frequency and $\beta=\frac{v}{c}$—velocity of electron. The fulfilment of the condition $\beta n>1$ is in this case definitely assumed, which implies the requirement that the velocity of the electron should exceed the phase velocity of light within the given medium. In reality, the motion of an electron in the medium involves a considerable change in velocity and, accordingly, magnitude β will vary greatly for different sections of the path of the electron. Hence, on the assumption that a but relatively small region of the spectrum is examined, within which n may be considered as constant, formula (1) with reference to the true motion of the electron should be written as follows:

$$W=\frac{e^2}{c^2}\int \omega d\omega \int\limits_{\beta n>1}\left(1-\frac{1}{\beta^2 n^2}\right) dl, \qquad (2)$$

where β depends upon l.

Integration along l will be accomplished with greater ease, if β and dl be respectively substituted by V and dV (V is the energy of the electron in kilovolts, dV—energy loss on section of path dl).

For such substitution we have in the first place the well-known relation $V=\frac{mc^2}{\sqrt{1-\beta^2}}-mc^2$; besides this, it follows from Madgwick's (3) experiments that the loss in energy of the electron per unit of path travelled within the range of energy, which we shall be concerned with in the subsequent (from about 200—300 kV up to 3000 kV), is for the given medium practically independent of V, i. e. $-\frac{dV}{dl}=k$, where k is constant. For different substances value k may with sufficiently good approximation be taken as proportional

to the density of the substance, provided only the atomic numbers of these substances do not differ markedly from one another.

Substituting in expression (2) value $1-\frac{1}{\left(\frac{V}{mc^2}+1\right)^2}$ for β^2 and $-\frac{dV}{k}$ for dl, we obtain:

$$W=\frac{e^2}{c^2}\int\omega d\omega\frac{1}{k}\int\frac{dV}{n^2\left[1-\frac{1}{\left(\frac{V}{mc^2}+1\right)^2}\right]},$$

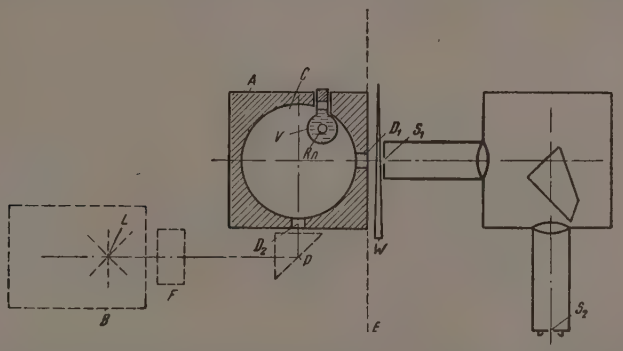
whence we have ultimately, upon carrying out the integration, for the total energy emitted by an electron within the range of frequencies from ν_1 to ν_2 the following expression:

$$W=\frac{2\pi^2e^2}{kc^2}\left(\nu_1^2-\nu_2^2\right)\left\{\frac{mc^2}{2n^2}\ln\frac{V_{\min}}{V_{\min}+2mc^2}+\left(1-\frac{1}{n^2}\right)(V-V_{\min})-\right.$$
$$\left.-\frac{mc^2}{2n^2}\ln\frac{V}{V+2mc^2}\right\}. \qquad (3)$$

Here V is the initial energy of the electron, $V_{\min}$—the energy of the electron corresponding to condition $\beta n=1$, below which the radiation is no longer called forth by the electron.

As we know the energy spectrum of electrons and the radiation of the source used, it will be possible from formula (3) to determine theoretically the expected magnitude of the absolute output of the radiation under consideration.

2. The arrangement by means of which the experimental determination of the output was effected is schematically represented in the figure. Heer



A is a massive lead block comprising spherical cavity C coated with magnesium oxide (Ulbrecht's sphere). The diameter of the cavity equals 13 cm. The liquid under investigation was placed inside this cavity in a separate thin-walled glass vessel V, assuming the shape of a small ball 3.5 cm in diameter. The radiation of the liquid was brought about by β-particles of Ra (B+C), having as their source radium emanation placed in a very thin-walled glass ampulla of spherical shape. This ampulla was placed inside the liquid in the centre of vessel V. Its diameter equalled 8 mm, while the thickness of the walls constitutes about 0.12 mm. In order to eliminate the luminescence of the glass of the ampulla walls, its outer surface was coated with silver.

Under such geometrical conditions the radiation of the source was found to be sufficiently homogeneous in all directions, though the emission of each

individual electron was strongly asymmetrical. The radiation of the liquid placed in vessel V proceeded practically only under the influence of β-particles since the thickness of the layer of the liquid in this vessel, being sufficiently great for the complete absorption of β-particles, was at the same time not great enough to cause any appreciable weakening of the hard γ-rays, emitted by the same source. The main portion of γ-radiation being absorbed by the walls of the lead block, the impedient action of these rays upon the eye of the observer was thereby obviated, as was also the possibility of luminescence of the glass parts of the arrangement.

Entrance slit S_1 of the monochromator with a breadth of 0.2 mm was placed in front of aperture D_1 made in the lead block. The breadth of the exit slit S_2 amounted to 0.91 mm. By means of this slit the region of the spectrum ranging from $\lambda=536$ mμ to $\lambda=556$ mμ was brought out, so that the mean wave length equalled 546 mμ (the wave length of the green line of the mercury arc). The radiation under investigation being extremely weak, it was found impossible to carry out direct measurements of the energy of the beam of light emitted from slit S_2. Accordingly, absolute energy measurements were accomplished with an auxiliary beam of light coming from a separate source L. As such a source the light of the green line of the mercury arc was selected. A beam from this light source, preliminarily filtered through combined liquid filter F [4], was directed through prism P of full internal reflection, then passed into aperture D_2 and finally reached spherical cavity C. By means of proper focussing (the optical system used for this purpose is not represented in the diagram) the intensity of the beam was adjusted in such a way as to make possible measurement with a thermopile. In these measurements the thermopile was placed inside spherical cavity C, in front of aperture D_2, in such a way that its working surface might completely overlap this aperture. The scale, from which the readings of the deviations of the mirror of the galvanometer were taken, was graduated in terms conforming to the general radiation of the Heffner lamp[5].

Further measurements consisted in the following. Coefficient a was determined after the quenching method[6], showing how many times the intensity of the beam discharged from the slit was to be lowered in order to bring this intensity to a value corresponding to the threshold of the visual excitation of the eye. This coefficient was consecutively determined for: 1) radiation produced in the liquid by β-particles and 2) the case when spherical cavity C was illuminated by an auxiliary beam of known intensity, W_1. Seeing that the albedo of the surface of spherical cavity C is close to unity while in other respects experimental geometrical conditions remain the same, and furthermore taking it for granted that the sensitivity of the observer's eye to light of a wave length of 546 mμ and the mean sensitivity within the range of wave lengths extending from 536 to 556 mμ may be considered as approximately equal, it will be possible from measurements 1) and 2) to find the absolute value of the output in question. Namely, it will be:

$$W = W_1 \frac{a}{a_1}, \qquad (4)$$

where the values with index 1 refer to the auxiliary beam.

In making observations according to the quenching method, the eye of the observer had its position directly outside slit S_2. The eye was shielded from outside light by screen E. The weakening of the light down to the threshold was effected by means of neutral-grey wedge W, the shifting of which was performed by the assistant. In the series of measurements with an auxiliary beam of light, besides wedge W grey filters were also provided for the additional weakening of light. Particular attention was paid to the most accurate determination of the constants of all the grey filters used.

3. The basic measurements of the absolute output of radiation were effected for water. Three series of experiments were carried out on different days. The amount of emanation used constituted: 193 mC in the first series, 162 mC in the second and 135 mC in the third. The intensity of the auxiliary beam was in each of the series equal to 1300, 1155 and 1120 ergs/sec. respectively. The final results of these experiments computed for one mC of emanation are given in Table 1.

Table 1

	Absolute output (per 1 mC of emanation) in ergs/sec.
1st series	$4.28 \cdot 10^{-4}$
2nd »	$3.91 \cdot 10^{-4}$
3rd »	$4.11 \cdot 10^{-4}$
Mean	$4.1 \cdot 10^{-4}$
Calculated	$3.5 \cdot 10^{-4}$

As seen from this table, the experimentally obtained value of the radiation output of water produced by β-particles of Ra (B+C) in the spectrum region extending from 536 to 556 mμ consitutes $4.1 \cdot 10^{-4}$ ergs/sec. per mC.

In the third column of Table 1 is to be found the value of the output computed according to formula (3), which after substituting numerical values for water ($n=1.334$, $V_{min}=260$ kV, $k=1710$ kV/cm) and for the region of spectrum under consideration assumes the following form:

$$W = 6.5 \cdot 10^{-14} \frac{1}{k} \left\{ 0.438\,(V-260) - 143.3 \ln \frac{V}{V+1020} - 228 \right\}.$$

In the process of calculation the whole energy spectrum of β-particles was divided into several sections, each of them being characterized by the number of particles and the mean energy of each particle. Data on the energy spectrum and the absolute number of β-particles emitted by Ra (B+C) were obtained from Gurney's(7) article. The data of the same author served also as a basis for the division of the spectrum of β-particles into separate sections.

The direct result of the calculation shows that within the range of the spectrum extending from 536 to 556 mμ the energy of radiation arising in water in the case of the passage of β-particles of Ra (B+C) through it equals $4.0 \cdot 10^{-4}$ ergs/sec. per one millicurie of Ra-emanation. Yet, on comparison with experimental evidence, this figure has to be somewhat reduced since, under the conditions of the experiments above described, part of the energy of β-particles is expended on the passage through the walls of the ampulla containing the emanation. The loss in energy occasioned by this circumstance equals for a separate β-particle 60—70 kV, which constitutes with respect to the mean energy of only those β-particles which satisfy the condition βn 1 a magnitude of the order of 12%. Since the passing through the walls of the ampulla takes place in the early part of the path of β-particles, it will be possible to assume approximately that the lowering of the output caused by this passage also constitutes a magnitude of the same order. Accordingly, the expected magnitude of absolute output should be $3.5 \cdot 10^{-4}$ ergs/sec. instead of $4.0 \cdot 10^{-4}$. This figure is recorded in Table 1. The magnitude obtained from experiments is, as has previously been stated, equal to $4.1 \cdot 10^{-4}$ ergs/sec., i. e. it is by 17% greater than the expected value. Making allowance for the specific conditions of the experiments (inconsiderable brightness, necessity of intermediate measurements of the intensity of the auxiliary beam of light, etc.), the aforesaid divergence may be considered as immaterial. Thus, we are free to assume that the experimentally obtained and theoretically calculated values of absolute output are in good agreement.

4. It is seen from formula (3) that with alteration of k and n, i. e. in passing over to liquids with other densities and refractive indices, the radiation output ought to be liable to change. In view of the fact that the distribution of radiation energy (over a limited section of the spectrum) is independent of n and k and is only a function of wave length ([8]), there is no need to effect anew any absolute measurements in order to determine the radiation output in some other liquid, but it will be sufficient with the availability of one and the same source of β-particles to evaluate by the integration method the brightness of the radiation of the liquid in question, in relation to the radiation brightness of another liquid (in the present case—water), for which the output has already been determined. Obviously, the output should be proportional to the brightnesses observed.

Similar measurements have been effected for several liquids (benzene, carbon disulphide, carbon tetrachloride, etc.). In these measurements, just as was the case in the earlier experiments, the ampulla with the emanation was placed into the liquid in the centre of vessel V (figure), while the vessel itself was located within spherical cavity C of the lead blocs. However, instead of a monochromator a photometer was installed before aperture D_1, this device being designed after Rayleigh's([9]) method. The results of the experiments are given in the following table:

Table 2

Liquid	Chemical formula of molecule	Density	Refract. index	$k = -\frac{dV}{dl}$ (in kV)	Output	
					Experim.	Calcul. after formula (3)
1	2	3	4	5	6	7
Water	H_2O	1.0	1.334	1710	1.0	1.0
Benzene	C_6H_6	0.879	1.505	1490	1.93	1.92
Cyclohexane	C_6H_{12}	0.787	1.433	1342	1.61	1.77
Carbon disulphide	CS_2	1.266	1.637	2165	1.68	1.70
Isobutyl alcohol	$C_4H_{10}O$	0.805	1.3985	1365	1.40	1.57
Carbon tetrachloride	CCl_4	1.61	1.468	2730	1.03	1.0

In the sixth column of this table the values of the output obtained experimentally are recorded, the output for water being taken as unity. The figures of the next column are computed according to formula (3). Here also, as the unit of output, the figure for water will be found to be taken ($4.0 \cdot 10^{-4}$ ergs/sec. per mC). In our calculations, the values $k = -\frac{dV}{dl}$ (5th column of Table 2) were in all cases taken proportional to the density of the liquid and were determined in the same way as for water from Madgwick's data for aluminium.

From a comparison of the figures of columns 6 and 7 in Table 2 it will be evident that for all liquids subjected to investigation, as well as for water, the values of the output experimentally obtained agree sufficiently well with those expected from theory. It seems of interest to note the fact that the energy expended by β-particles of Ra (B+C) on the excitation of the radiation under consideration constitutes a value of the order of only 0.1% of the total energy of these particles.

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PHYSICS

HEAVY ELECTRONS IN COSMIC RAYS

By V. VEKSLER, K. ALEKSEEVA and N. REYNOV

(Communicated by S. I. Wawilow, Member of the Academy, 23. IX.1938)

The experiments effected by Veksler and Isajev(1) have shown strongly absorbed particles to be present in the composition of cosmic radiation. Veksler and Dobrotin(2) found those particles to be heavy electrons. They have, moreover, determined the ranges of the values of heavy electron mass as well as the dependence of their number on the thickness of the lead plate placed above the arrangement.

A further study of the properties of heavy electrons was conducted by us during the mount Elbrouz expedition in 1938.

For this work we had recourse to the method of proportional counters, already described by us in our previous articles. The parameters of the proportional counters and that of the amplifier were identical with the previous; we will, therefore, not give them in the present paper, which contains the results of these investigations.

Description of Experiments. The results of our preceding works on strongly absorbed particles may be interpreted not only if we assume the presence of heavy electrons, but also if we admit that the arrangement made use of was registering groups consisting of usual slow electrons. But since doing so one would be bound to assume that the number of such groups of slow electrons becomes considerably reduced when small thicknesses of lead are placed above the arrangement, such an assumption seems scarcely probable. Yet, to obtain definite proof of that our arrangement was actually registering individual strongly ionized particles (in this case alone the evaluation of their mass, suggested by Veksler and Dobrotin, holds true), we arranged experiments with triple coincidences.

The position of the proportional counters and their switching on the scheme of triple coincidences are shown in Fig. 1. Lead plates of various thicknesses were placed above the upper group of counters. The number of triple coincidences was counted, a thin aluminium plate (0.55 g/cm^2) having been put between the upper and two lower groups of counters, and without it. The results of these measurements are given in Fig. 2. Along the axis of the abscissae is plotted the thickness of the lead plate, placed above the counters, whereas along the axis of ordinates at the right—the per cent of coincidences, eliminated by the aluminium plate, at left—the number of non-eliminated coincidences (showers of usual electrons).

Curve *2* represents the Rossi curve for showers on triple coincidences (those non-eliminated by the aluminium plate). By circles(Curve *1*) are denoted the values of the per cent of coincidences, eliminated by the aluminium

plate at different thicknesses of the lead plate, placed above the counters. For the sake of comparison we give in the same figure Curve *3*, representing the per cent of coincidences eliminated by the aluminium plate in the scheme of double coincidences*. From this figure it may be seen that in the case of double coincidences the effect of the aluminium interlayer considerably exceeds that observed in the case of triple coincidences. It means that double coincidences eliminated by the aluminium interlayer cannot be produced by groups of particles, but should be attributed to the action of heavy electrons. Hence it follows that a considerable portion of double coincidences is in general being produced by separate strongly ionizing heavy electrons. On this ground, in our investigation of the dependence of the number of heavy electrons upon the altitude we had recourse to the scheme of double coincidences.

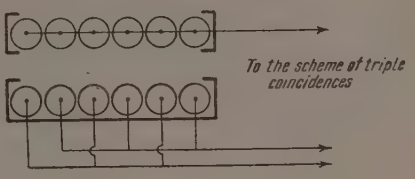

Fig. 1.

Dependence on Altitude. To determine the dependence of the number of heavy electrons upon the altitude, measurements have been effected at three different points: Krugozor (3000 m above sea level), Priyut-9 (4200 m) and Sedlovina of mount Elbrouz (5300 m). The results obtained are given in Fig. 3. The experimental dependence of the number of showers (double coincidences non-eliminated by the aluminium interlayer) upon the altitude, found by us, is plotted

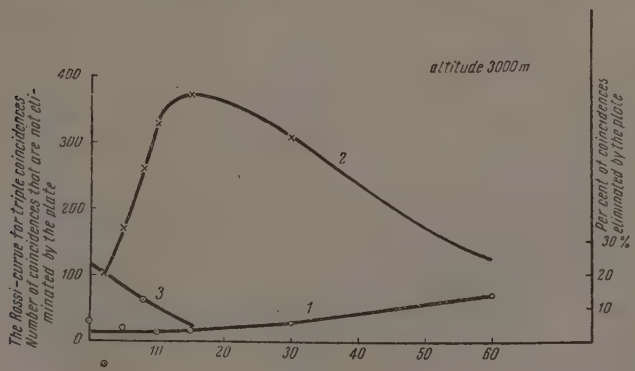

Fig. 2.

on Curve *1*. The number of heavy electrons observed by us at similar altitudes (coincidences eliminated by the interlayer) are denoted by circles.

For the sake of comparison the same figure contains curves [plotted according to Heitler's data(3)] which illustrate the growth of the soft and penetrating** component of cosmic radiation (Curves *2* and *3* respectively).

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* To keep geometrical conditions unchanged when passing from triple coincidences to double ones, in the experiments with the latter we made use of all the counters of the upper box (5 counters), and one branch of the lower one (3 counters) (Fig. 1).

** Curve *3* is drawn across the point which corresponds to the number of heavy electrons at sea level.

From Fig. 3 it may be seen that the dependence of the number of heavy electrons upon the altitude coincides with the dependence upon the altitude of the hard component of cosmic rays.

Discussion of Results. A comparison of Curves *1* and *3* in Fig. 2 and the dependence of the number of double coincidences eliminated by the aluminium interlayer upon the altitude shows that the appearance of strongly ionizing particles is not connected with the formation of showers and that at least 75—80% of double coincidences, which may be eliminated with the help of an interlayer, are due either to single heavy electrons, or to heavy electrons, accompanied by a very small number of usual ones.

As to the triple coincidences eliminated by means of an aluminium interlayer, they may be accounted for by groups (showers) of heavy electrons.

Although the assumption of usual slow electrons being responsible for those coincidences cannot be definitively refuted, yet such an assumption would contradict to the fact that the percentage of triple coincidences under lead does not increase (which should take place owing to a considerably lower critical energy of lead as compared to that of the air), but even tends to diminish.

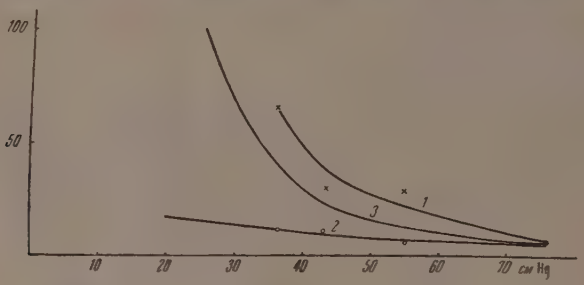

Fig. 3.

A comparison of the data obtained by us with the results of our previous experiments enables us to advance some assumptions as to the formation of heavy electrons.

Fig. 3 shows that heavy electrons are in equilibrium with the penetrating component of the cosmic radiation. Yet, it had already been shown[2] that those heavy electrons are secondaries (or tertiaries, and so on), and may thus not be regarded as the «tips» of paths of particles entering into the primary penetrating component. Hence it may be derived that they are produced by the penetrating component either directly, or through some intermediate mechanism. An analysis of the data as to the dependence of the number of heavy electrons on the thickness of the lead plate placed above the counters may supply some information in this direction (Fig. 4)*. A comparison of the rate of increase of the number of heavy electrons with altitude (Curve *2*, Fig. 3) with the sharp reduction of their number when small thicknesses of lead are placed above the counters, points to that this drop cannot be accounted for by the shape of the energy spectrum of the given particles, but should be rather regarded as a result of the transition effect from air to lead.

We may thus assume that in this case we have two different processes. On the one hand, there takes place a reduction of the number of particles

* The curve is taken from the article by Veksler and Dobrotin.

formed in the air due to their absorption by lead, on the other hand, new heavy electrons are formed in lead.

Assuming the first of the two processes mentioned above to prevail in the initial part of the curve (Fig. 4), one may roughly estimate the paths of the particles formed in the air, and calculate the number n of heavy electrons that should be in equilibrium with a single penetrating particle. Such an evaluation shows that n must be near to 0.2—0.4. It is clear enough that such a high value of n cannot be conciliated with the experimental data (in particular with observations effected by means of the Wilson cloud chamber), if it be considered that penetrating particles produce heavy electrons directly or through secondary ionizing particles.

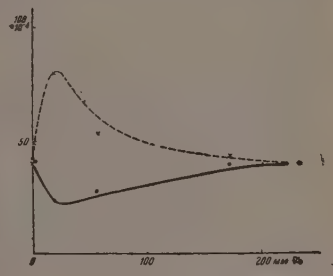

Fig 4.

This discrepancy may be, obviously, done away with if we assume that heavy electrons are produced by penetrating particles not directly, but with the help of an intermediate non-ionizing component. However, this non-ionizing radiation cannot consist of photons, since the latter are contained by the soft component of the cosmic radiation, while heavy electrons have been determined by us to be in equilibrium with penetrating particles.

We thus arrive at the hypothesis that there must exist some non-ionizing particles which are produced by the penetrating component and which are themselves capable of producing heavy electrons. Not contradictory to the experimental data available, this hypothesis can be used to interpret some other phenomena observed. Indeed, the sum of data concerning the second maximum of the Rossi curve (range of shower particles, position of the maximum, dependence on Z, etc.) seem to exclude the possibility of the second maximum being created directly by the penetrating particles or photons. On the contrary, the increase of the number of heavy electrons in the case of thicknesses corresponding to the position of the second maximum of the Rossi curve, found by Veksler and Dobrotin[2], may be placed in connection with this phenomenon. The same standpoint may be applied to explain the observations of Wilson[4], who found that the particles of the penetrating component suffer considerable energy losses which are accompanied neither by photon, nor by ionizing particle production.

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BIOGEOCHEMISTRY

BIOCHEMICAL RÔLE OF ALUMINIUM AND SILICON IN SOILS

By W. I. VERNADSKY, Member of the Academy

A. A. Rode published recently in the journal «Pochvovedenie» (No 2, 1938) a most interesting and noteworthy article on the dispersity of soils. This article includes (see page 214) a list—unfortunately far from being complete—of data as to the mineralogical composition of the finest particles of soil hitherto considered by pedologists as colloidal. It is a matter of regret that, judging by his paper, the author as well as the other pedologists of this country fail to realize all the new and important conclusions that can be derived from this table.

All the crystalline minerals identified by X-ray analysis the author refers to belong in one definite group of minerals, namely, to the kaolinic or alumosilicic hydroxyacids.

This fact has previously been stated for soils using another method, viz. that of thermal analysis.

It can, thus, hardly be any doubt as to the existence of this phenomenon observed by two independent and accurate methods, nor do we possess any data indicative of the formation of these minerals outside the soil.

It follows from this fact that, as far as we can judge, everywhere in soils is going on a process of new formation of free crystalline kaolinic alumosilicic acids (kaolinic clays).

As I was aware that the formation in soils of free alumosilicic acids of the kaolin range, which was stated in the last years, has an enormous importance for the comprehension of the phenomena of plants' nutrition and the biochemistry of aluminium and silicon, I proceeded to a special investigation of this problem at the Biogeochemical Laboratory of the Academy of Sciences, after the latter was transferred to Moscow in 1935. I started from the empirical generalization that the process in question is influenced by the living world of the soil being a biogeochemical, fundamental process hitherto unknown in the geochemistry of aluminium and silicon.

The results of this work which is as yet not achieved show clearly that this generalization has been true.

Another interpretation of the above phenomena, essentially differing from what is being exposed here, may be and has actually been suggested by A. A. Rode. I started from the results obtained by French pedologists and mineralogists which seem unfortunately to have remained unknown to A. A. Rode.

Prof. W. C. Agafonov was the first to apply the thermal analysis to soils and the methods of V. V. Dokuchaev to the study of soils of France and Tunis. W. C. Agafonov showed the wide distribution of the crystalline

kaolinic acids in the finest soil particles formerly considered as colloidal. Now we know that his conclusion holds true for all soils including this country. It should be remembered that the kaolinic alumosilicic acids—montmorillonite, halloisite, kaolinite, dickite, and so on, which are formed in the soil, have not been subjected to a thorough mineralogical investigation. The process of their formation is very complicated and the precise formula not always quite clear. This applies especially to the most common of them, the montmorillonite. It is similar to the hydrate $H_2Al_2Si_4O_{12}\ n$ aq., but always contains Mg; thus, we can suppose that it is followed by disintegration of magnesia alumosilicates (micas). It will be necessary to reproduce the montmorillonite in a pure form synthetically and to study the products of the bacterial decomposition of micas.

Biogeochemically we may regard these new formed crystalline kaolinic alumosilicic acids as a result of bacterial and, possibly, fungous processes, of the microbial fauna of the soil and its feeding on potassium derived from the salts of kaolinic alumosilicic acids (feldspars, micas, and so on), always present in detritus of soil. This phenomenon is common to all soils, so that this biochemical new formation of the kaolinic aminoacids must be taken as a process common to the geochemistry of aluminium and silicon.

As a matter of fact, the question as to the absorptive capacity of soils wants to be revised, and to a certain extent we may find ourselves obliged to come back to the old interpretation.

We should not forget that the soils, overfilled with living organisms as they are, present anything but inert natural bodies. Being regularly built up from both living and inert material, they should rather be called «bioinert» natural bodies of the biosphere. When absorption phenomena in soils were discussed in soil-science, one generally took into account the chemical processes and the manifestation of partial forces while the greatest force present in soil, «the living matter»*, was left out of the question. The scientists in this country have of late attributed a particular importance to the action of molecular forces (colloid processes); they fail, however, to give a satisfactorily clear picture of the process. In the case of Gedroyc's «absorbing complex», for instance, the action of molecular forces may not account alone for the separation of aluminium from silicon in a kaolinic nucleus. In my private discussions with the late C. A. Gedroyc I pointed out that the kaolinic nucleus $Al_2Si_2O_7$ is an endothermical compound, scarcely decomposable even by strong chemical acids. Yet, as we know by experience, it can be easily decomposed biochemically.

So great is the rôle of the living matter in soils that no experimental results can be taken into account, if the experiments were conducted in presence of life and not under sterile conditions. The living matter manifests itself in such an experiment within a few hours, if the conditions are favourable to life. This must be borne in mind while experimenting with the absorption of salts by soils; sterile conditions for such experiments are indispensable. For ought I know, no such experiments have been carried out.

Soil absorption can be either a biological dynamical process of bacteria-life, or a chemical process—formation of acid salts from free kaolinic alumosilicic acids produced bacterially, or else a manifestation of molecular forces, except for the case of kaolinic nucleus.

The responsible cause will remain a matter of controversy until after new experiments have been conducted along new lines. The more so as the question of the rôle played by molecular forces in the salt absorption of soil needs reinvestigation.

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* Living matter corresponds to the ensemble of living organisms.

We probably have to deal with a complex and peculiar biogeochemical process—bacterial formation of kaolinic acids from kaolinic salts (feldspar, mica). Our experiments, which will be soon published, show that very clearly. It should also be kept in mind that alongside of bacterial formation of free alumosilicic kaolinic acids, we find a more profound process to break up the kaolinic nucleus in the soil—the separation of aluminium from silicon by another biochemical process, by the action of diatoms always present in soils. This phenomenon was discovered in oceanic waters by Murray. Experiments carried on in our laboratory have established that a pure culture of diatoms will break up the kaolinic nucleus, but the development of diatoms is better and their reproduction more rapid in presence of bacteria.

The question of the absorptive capacity of soils is of such an importance for the problem of agriculture, that it requires a more extensive and close study and revision.

This was the reason why I thought it expedient in the present article to outline the future consequences of our work. Pedological and agricultural laboratories which are better equipped for such experiments should immediately take joint action on our research.

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PETROGRAPHY

SOME QUERIES ON CLASSIFICATION AND NOMENCLATURE OF IGNEOUS ROCKS. II*

By F. LOEWINSON-LESSING, Member of the Academy

The classification and nomenclature of igneous rocks is still characterized by some vagueness and wide differences of opinion. This classification and nomenclature, which is a product of historical growth, involves inconsistencies, controversial matters, a lack of agreement between various petrographic schools. All attempts to effect a radical reconstruction, to find new ways and to bring about some kind of agreement have so far been futile. It is just possible that the «Petrographic Association», planned at the International Geological Congress in Moscow, may in time succeed in bringing about some kind of harmony and in conciliating divergent views. In my opinion, no hope of an immediate agreement can, however, be held out. Nevertheless, some controversial points in the classification and nomenclature of igneous rocks are not only of interest to the theoretical scientist, but compel the attention of the lecturer and the compiler of text books. This is why, without tackling the whole of the problem straight away, I have adopted the plan of publishing periodically, as I did it at the end of the last century and at the beginning of the present one (2), a series of short articles dealing critically with some particular questions that are of special interest to me and stating my point of view in regard to them.

In the first place, I shall consider some questions connected with the problem of the parallelism of intrusive and effusive rocks, a problem I have dealt with already elsewhere.

1. Andesites. I have repeatedly considered the family of andesites both from the point of view of its subdivisions and from that of it delimitation from the family of basalts and dacites. The matter we have to consider at present is of a different nature, namely the systematic position of andesites in the parallelization scheme of the families of effusive and intrusive rocks. On the background of qualitative mineralogical classifications a parallelism was established between andesites and diorites, i. e. the generally adopted view was that andesites are the effusive facies of diorites, and vice versa. If, in qualitative mineralogical classifications such a parallelization does not seem to give rise to any doubt, it has long been known that there is an essential difference between the chemical aspect of andesites and diorites. This difference consists not only in a higher content of silica (also observed as a rule in other parallel series of intrusive and effusive rocks), but also in the alkali content, in the coefficient of acidity and in the aspect of the magmatic formula. Properly speaking, andesites

* The first of this series is an article published as far back as 1928 (1).

are not basic but neutral rocks, saturated with silica and, as regards their chemical aspect, much more closely related to syenite than to diorite. The only point on which andesites differ from syenites is practically their R_2O : RO ratio, and, more particularly, a lower content not only of the total amount of alkalis, but more especially of potash. This accounts for the absence of potash feldspar, which is invariably an essential and characteristic component of most syenites. As already stated by the writer in his «Petrography», the felsophyrite described by Slavik and Orlov is, according to its chemical aspect, a true effusive equivalent of diorites.

In establishing a parallelism between effusive and intrusive rocks and in approaching the problem of an independent magmatic origin of some families of igneous rocks, it is, in my opinion, a matter of importance also to take into account the quantitative ratio of that family to the total composition of the earth's crust. Those rocks that are widely distributed and occur as large and numerous intrusive bodies or as vast flood extrusions may be regarded as due to a direct crystallization of a particular melt, of a particular magma. A still more definite proof of the existence of such a particular magma is furnished by such instances where a certain chemical type is represented by both flood extrusions and numerous large intrusions. This condition is fully satisfied by basalts with the gabbroic or a doleritic formation, and by quartz porphyries with a granitic formation. This condition, however, is far from being satisfied by diorites. The latter do not, generally speaking, show a very wide distribution, and, in measure as the term «diorite» gains in presicion and various pseudodiorites, paradiorites, metadiorites, epidiorites etc. become differentiated, the quantitative coefficient of the proportion of diorites in the composition of the earth's crust is reduced.

Andesites are not true basic rocks, but neutral rocks saturated with silica; still, they are rocks derived from a basic magma. Herein lies the essential difference between andesites and basalts: the latter are products of a direct crystallization of the basic magma; the former are products of some kind of a modifying process (assimilation? differentiation?) that took place in the basic magma. This point of view was adopted by Daly ([3]), and it seems to me that he is right in this respect (but andesites are nevertheless referred by him to the family of diorites). This would account for the fact that andesites, although widely distributed, do not form, as far as I know, any fissure eruptions, being only effusions from central volcanic vents. We must therefore conclude that andesites are a peculiar effusive type, derived from a basic magma and having no analogue amongst intrusive rocks. Rocks of this type may be termed a p o m a g m a t i c effusions as distinct from o r t h o m a g m a t i c ones, i. e. those represented by chemically identical effusives and intrusives, as, for instance, basalts and gabbro.

2. T r a c h y t e s. A problem of classification and parallelization similar to that raised by andesites also arises in connexion with trachytes. In mineralogical classifications, trachytes are parallelized with syenites. A certain difference, however, is observable even in their mineralogical composition, since plagioclase, an essential component of syenites (except the alkaline syenites group), is either absent in trachytes or plays a subordinate part. The alkaline nature of trachytes is unmistakably reflected in their chemical aspect which clearly suggests that trachytes are an effusive form of a less comprehensive group of intrusives, viz., of alkaline syenites, and not of syenites generally. A good illustration of this may be found in the comparative table given below.

This being the case, one is led to conclude that ordinary syenites have no effusive equivalent; in any case, neither trachytes nor andesites are

such equivalents. This fact is very important. To my mind, it points to the fact that there is no such thing as a special syenitic magma, that syenites are not derived from some special magma of syenitic composition, but have been derived from either the acid or the basic magma by processes of assimilation or differentiation which have not yet been fully ascertained. Syenites are therefore a p o m a g m a t i c intrusions as distinct from o r t h o m a g m a t i c ones, such as granites and gabbro.

	$\overline{RO} : SiO_2$	$R_2O : RO$	α	SiO_2, %
Alkaline syenites	1.5 : 5.75	1.56 : 1	2.55	60.78
Trachytes	1.27 : 5.31	1.26 : 1	2.48	60.68
Syenites	1.8 : 5.6	1 : 2.2	2.34	60.19
Andesites	1.48 : 5.21	1 : 2.6	2.32	59.59
Diorites	1.5 : 4.3	1 : 4.3	1.77	55.46

Daly (4) also regards syenites as derived (i. e. according to my terminology as apomagmatic) rocks, and not as products of a direct crystallization of a neutral magma.

In approaching the problem of the origin and the genetic connexion of syenites one should keep in mind the dual nature of syenites. On the one hand, syenites are connected both geologically and by successive links of mineralogical and chemical composition with the gabbro formation, more particularly with that type of the latter which I have termed the Ukrainian or Norwegian type. On the other hand, syenites are similarly associated with granites. It should be kept in mind, however, that these are two different branches, two distinct types of syenites, and that there are in existence no such intrusive formations as would display a continuous sequence of contemporaneous transitions from gabbro to granite. No comparative characterization of the two branches of syenite rocks, the gabbroic and the granitic one, has so far been given. It would appear, however, that both chemically and mineralogically it might be summarized as follows: the gabbro branch of syenites has a comparatively higher content of alkaline earths, plagioclase being more prominent in it; in the granite branch of syenites the ratio is reversed: it shows a tendency to a certain supersaturation with silica which accounts for the occurrence of sporadic quartz.

Alkaline syenites are genetically associated with the nepheline-syenite formation, in the same way as trachytes, their effusive facies, are associated with phonolites, the effusive facies of nepheline-syenites.

The table given above shows the main elements of the chemical composition of the rocks in question. The essential difference between andesites and diorites is clearly brought out by this juxtaposition. Andesites are closely related to syenites, the only essential difference being the value of the $\overline{RO}$ coefficient, i. e. their more leucocratic nature, and also the alkali content; although the $R_2O : RO$ ratio is about the same in both syenites and andesites, still the former have a higher absolute alkali content and a higher percentage of potash in the total amount of alkalis. This accounts for the difference in feldspars. The ratios of the orthoclase particle to the albite and anorthite particles, as calculated from chemical analyses of syenites and andesites, are (on average):

for syenites —Or : Ab : An = 2 : 3 : 2
for andesites—Or : Ab : An = 2 : 5.5 : 8

I raised this question already as early as 32 years ago (5). The answer elicited by it in those days was: «There are no essential differences in the

chemical composition of effusive and intrusive rocks»... «On the whole, the same chemical types recur in both intrusive and effusive rocks, although the frequency of occurrence of the same type as an intrusive and as an effusive rock and its significance in the composition of the earth's crust are often different»... «Effusive rocks display a great variety of details in their chemical composition and show a greater number of transition types characterized by comparatively slight differences in chemical composition». In that paper, andesito-basalts, not andesites, were classed by me as effusive analogues of diorites. As to andesites, they were parallelized with syenites, and trachytes with åkerites (i. e. with alkaline syenites).

3. Spilites. A great deal has been written about spilites ,and their aspect, both chemical and geological, would seem to be fairly well defined. It may be considered as proved that the most striking peculiarity of spilites is not to be found in their structure, as formerly believed, but in their chemism (their high soda content), their mineralogical composition (primary albite or albitized basic plagioclase), in their geological characteristics (extrusion into water), and finally in the texture of spilitic lavas (pillow lava). None of these peculiarities is sufficient if taken singly, but taken as a whole they are characteristic enough. Contrary to some opinions, spilites are not an intrusive, but an effusive (and subintrusive) facies of the basic magma, which was cooled under water, and not on land. Finally, the geological association of spilites with keratophyres, emphasized by Dewey and Flett, lends them a peculiar character and is to some extent analogous to the association of pacific basalts with trachytes.

Some discrepancies may occur within the limits of the above definition, but it may on the whole be regarded as a widely accepted one. Serious differences and doubts arise, however, when we approach the essential genetic problem of the origin of the soda which is the most characteristic peculiarity of spilites (and of the accompanying keratophyres). This, i. e. the soda and albite content, is practically the only essential point of difference between spilites and basalts.

As to the origin of the soda of spilites, two alternative explanations suggest themselves: 1) the surplus of soda as compared with basalts is inherent to the spilitic magma itself; 2) this surplus soda is derived from outside in the process of solidification and crystallization of the spilitic melt. The ground for the latter point of view may be found in the fact that in many instances, and, for quite a number of authors, in a great majority of cases, spilites are not albitic, but albitized rocks. In other words, the first to be segregated was a basic plagioclase which subsequently, by a process of protopneumotolysis, became albitized at the expense of some (most probably, volatile) compounds of soda. In view of the fact that the solidification of spilites took place under water, one might have looked for the source of soda in sea water. Two facts, however, are in my opinion inconsistent with this assumption. In the first place, I do not see how the salts of sea water can produce albitization. The sodium chloride of sea water might produce scapolitization, but it could not have produced albitization; basic plagioclase cannot be transformed into albite except in the presence of sodium silicate, or in case of the combination of sodium carbonate and silica. In the second place if the soda for albitization were furnished by sea water, any submarine extrusion of basaltic magma ought to yield spilites. We are aware, however, of submarine diabase extrusions (diabases, porphyrites) with protopneumatolytic chlorite, but not albitized. In other words, basaltic magma, if solidified in a sea basin, yields, in some instances, a spilitic facies, and, in some others, a diabasic facies; this would be unaccountable if the source of the soda needed for the albitization were sea water; it would,

on the contrary, be quite natural, if the spilitic facies were produced when the basaltic magma carried with it some particular compounds that are necessary to bring about its albitization.

Thus, there remains the other alternative: the spilitic facies is yielded by basaltic magma when the latter is accompanied by such compounds of soda (and by silica?) as are necessary for that combination; in other words, the basaltic magma is not homogeneous, being in some cases primarily enriched with soda. This assumption would also find a confirmation in the intrusive facies of basaltic magma. We are well aware that sodic contacts (adinoles, spilosites, desmosites), which are characteristic of basic magma, as distinct from acid magma, are not necessarily present in a gabbro formation. Side by side with such contact modifications, gabbros also yield ordinary hornstone contacts. Similarly, there are known intrusive gabbro-diabases containing primary albite, while no explanation of this fact may be found in their environment.

I suggest as an objective inference from the above considerations that the spilitic series of basaltic magma is a facies primarily enriched in soda, and that soda accompanies basaltic magma as gaseous or fluid compounds, which take part in the crystallization or the magmatic metasomatization of the basaltic effusive in course of crystallization. Similarly to this, the intrusive facies, i. e. the gabbro formation, gives rise in some cases to ordinary hornstone contacts, and in others to soda-albitic contact modifications. The cause of this is unknown to us, and we have to accept as a fact the lack of primary homogeneity in basaltic magma from its very origin, as we have other facts of the same kind, such the heterogeneity of granitic magma (why are not all granites stanniferous?), that of ultrabasic rocks (why are not all dunites, even if only in the Urals, platiniferous?), and so on, and so forth.

It should, however, be kept in mind that ordinary basalts are an usual type of basic effusives, while spilitic basalt is, in spite of its frequent occurrence, only a subordinate type, alien to continental fissure eruptions and to intrusive sheets.

The spilitic facies of basaltic magma would appear to be most frequently associated with geosynclinal regions, the connexion being, however, by no means so universal and unalterable as asserted by some geologists and petrographers. Not all geosynclinal regions are characterized by the spilitic formation; it would suffice to mention the Urals with their diabasic and porphyric formation, where spilites either play a subordinate part, or are altogether absent.

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STRATIGRAPHY

NEW DATA ON THE STRATIGRAPHY OF UPPER JURASSIC AND NEOCOMIAN OF NORTH-EASTERN AZERBAIDJAN

By N. B. VASSOEVICH

(Communicated by A. A. Borissiak, Member of the Academy, 8. IX.1938)

The first scheme of the stratigraphy of Cretaceous and Jurassic deposits of south-eastern end of Main Caucasian Range was suggested by K. I. Bogdanovich in 1902—1904. This scheme, published in the monograph «System of Dibrar» ([1]), has such an aspect:

Khaltan series Flysh facies of Tithonian and Neocomian	7. Consolidated red-brown sandstones 6. Light-grey glauconitic sandstones 5. Sandstones and clays with inclusions of spherosiderite and interbedding marl with cone-in-cone structure
Cenomanian	4. Horizon with *Phylloceras forbesianum* d'Orb. 3. Clays of Terebratulina series
Lower Senonian and Upper Turonian	2. Horizon with *Actinocamax*
Orbitoidal series Upper Senonian	1c. Dark-grey marlaceous clays interbedded with calcareous sandstone or brecciated limestone 1b. Red-coloured marlaceous clays and white marls with *Fucoides* 1a. Fine schistous marlaceous clays and consolidated marls of white, light-grey and greenish colour, with *Fucoides* and hieroglyphs

A first important correction to the scheme of Bogdanovich in 1908 was proposed by Stolley who demonstrated the Aptian age of the horizon with *Actinocamax*.

The «Orbitoidal strata» were subjected to revision by N. S. Shatski and M. F. Mirchink. They have found that the Orbitoidal strata embrace a whole series of different chalk horizons, bedding over the Aptian.

In 1931 Z. A. Mishunina determined anew the fauna of upper part of Terebratulina series of Bogdanovich and came to the conclusion about the Lower Aptian and Upper Barremian age of the upper part of «Terebratulina beds». This author also established that the beds «3» and «4» of Bogdanovich' scheme were nothing but facies of the same series([7]).

Towards 1935—1938, as a result of investigations of a number of geologists, particularly I. M. Goubkin, M. F. Mirchink, Z. A. Mishunina, N. S. Shatski and others, it became possible to develop new schemes of cretaceous stratigraphy, published in the works by M. F. Mirchink([5,6]), I. M. Goubkin([3]), V. E. Khain([8]), I. M. Goubkin and S. F. Fedorov([4]). These sections are given in the table.

In late autumn 1937 and in the beginning 1938, in behalf of the Azerbaidjan Institute for Oil Research I made a geological exploration of the north-eastern part of the ASSR. The object of these explorations was to make more precise the stratigraphy of cretaceous deposits. After a detailed study of sections of Haltan series and of more young series in the Dibrar district, Haltan thermal sources and in the by-Caspian oil-bearing region, we succeeded in developing a more complete stratigraphic scheme, a brief description of which is the subject of this note.

Gilginchai series. The more ancient deposits protruding in the cores of anticlinal folds—Khaltan, Kizil-Kazma and Besh-Barmak, are represented as dark argillaceous shales with interbedding sandstones. This thick series should be referred to the Dogger age.

Kizil-Kazma (Khaltan s. str.) series. Over the Gilginchai strata there occurs a discordant thick series of sandstones, usually slightly calcareous, with interbedding alevropelites and separate beds of conglomerates. The colour of the rocks is variegated: the brownish-red sandstones are developed side by side with greenish and grey ones. In conglomerates the boulders occur both of light zoogenic limestones of Besh-Barmak type (Lusitanian series), and of dark bituminous limestones.

The regional-stratigraphic correlations allow to refer the latter to the Oxfordian (the Oxfordian limestones are to a considerable extent bituminous in the zone of the north slope of Great Caucasus in the basin of Assa and Terek rivers, in Mountainous Ossetia, etc.).

The general lithological character of Kizil-Kazma series (Khaltan s. str.) is common as well to the shallow deposits of Kimmeridge and Lower Tithonian described by L. A. Vardanianz for the district of the Ossetia Military Road, and to the basin of Fial-Don river, as to the variegated flysh thickness of Upper Jurassic, developed on Pchekha river (n.-w. Caucasus) and investigated in detail by V. V. Belousov.

Taking into account the confusion in the understanding of the volume of «Khaltan series», it is necessary to remember that Bogdanovich, who was the first to separate it, in the first lines of his «System of Dibrar», confined to the description of the new series of rocks, states with sufficient explicitness: «For the first time in the deep sections of the Kizil-Kazma river appear the beds of new series, which north-westwards gradually assume larger dimensions. This series of beds which has the form of more or less calcareous sandstones of grey, dirty-greenish and brownish colour, I will nameKhaltan».

K. I. Bogdanovich emphasized too that between the Khaltan series and the overlying beds there is a sharply expressed break, and where an «apparent conformability is observed, there occurs a whole range of other formations». Thus, as a matter of fact, one can determine as Khaltan series only Kizil-Kazma sandstones, the thickness of which grows north-westward at the expense of the Upper Tithonian denudation up to 860 m.

The general lithological habitus of the Kizil-Kazma series, the presence in the conglomerates of shoulders of Lusitanian and Oxfordian limestones, the unconformity in the base of series and the transgressive overbedding of the Nardaran series,—point to the age of the series, as referable to the Kimmeridge—Lower Tithonian.

Nardaran series. Upon the Kizil-Kazma series, or sometimes even directly upon the J_{2-1}, lies with a more or less expressed occasionally dip-unconformity a series that had been referred by Bogdanovich sometimes to the Orbitoidal (Nardaran gully), sometimes to the Terebratulina beds (Kizil-Kazma river), or to the Khaltan series (Khata-dere river). This series, which upon field examinations in November 1937 were given the name of «Nardaran», has a considerable facial variability. In the Khaltan district it is represented by flysh; northward from the Gilginchai river—by calcareous marlaceous complexes of rocks. In the neighbourhood of Besh-Barmak the series has an intermediate character.

Columa of Upper Jurassic and Lower Cretaceous Beds of South-Eastern Azerbaidjan from Data of Different Investigators

K. Bogdanovich, 1906		M. F. Mirchink, 1931 and 1935		V. E. Khain (after M. F. Mirchink, Z. A. Mishunina and V. E. Khain), 1937		I. M. Goubkin, 1934, I. M. Goubkin and S. F. Fedorov, 1938		N. B. Wassoevich, 1938		
Formation (local subdivisions)	Age	Formation (local subdivisions)	Age	Formation (local subdivisions)	Age	Formation (local subdivisions)	Age	Formation (local subdivisions)	Index	Age
Orbitoidal series (lower portion of the horizon)	Upper Senonian	Subkemchi series: Kululu horizon	Cenomanian	Subkemchi series	Lower Turonian, Cenomanian and Upper Albian	Subkemchi series	Cenomanian	*Aucellina gryphaeoides* horizon	Kl_2	Upper Albian
		?		? ? ?		?		Kululu sandstones (in some parts unconformity)	Kl_1	
Discontinuity		Discontinuity	Albian	? (Unconformity and discontinuity)	Lower Albian	Discontinuity	Albian	Tankalasina horizon	Tnk	Middle and Lower Albian
		?		?		?		Alty-Agatch series	Alt	
Actinocamax horizon	Contact line of Senonian and Turonian	Khanagla series In some parts unconformity	Aptian	Red-clays series	Upper Aptian and upper part of Lower Aptian	Red-coloured series	Aptian	Khanagla series	Hn	Aptian

Beds with *Phylloceras forbesianum* d'Orb. and the Terebratulina beds		Cenomanian											
Discontinuity													
Khalian series (sensu lato)	Sandstones and clays with spheroside-rite	Neocomian and Tithonian	Series of grey clays (Khalchai series)	Aptian-Barremian	Grey-clays series	Barremian and lower part of Lower Aptian	Terebratulina horizon	Aptian-Barremian	Khalchai series	Hal	Lower Aptian (?). Barremian
							Phylloceras horizon	Valanginian-Hauterivian	Kaytar series	Klr	Hauterivian
	Light-grey sandstones with glauconite		Beds *Phylloceras*	Neocomian	Khaltan series. In upper part light-grey limestones and marls, in lower part sandstones	Neocomian-Tithonian	Khaltan series		Nardaran series	Nrd	Valanginian and Upper Tithonian
			?		?						
	Khaltan series (s. str.). Sandstones of the river Kizil-Kazma		Khaltan series (Discontinuity inside «series» denied. *N. V.*)				(facies of the Nardaran series. *N. V.*)		ordinarily unconformity		
							Shah-dag limestones	Tithonian	Kizil-Kazma series (Khaltan series s. str.)	Kz	Lower Tithonian and Kimmeridgian
									Unconformity		
									Gilgiuchai series	Glg	Dogger

At the base of the Nardaran series are of frequent occurrence conglomerates, reaching a thickness of 250 m (south wing of Khaltan anticlinal). In the composition of conglomerates enter the rocks of all the underlying series from J_{2-1} and upwards. In some individual cases the conglomerates are absent, ant the transition to Kizil-Kazma (Khaltan s. str.) series is rather imperceptible, for instance, at the north wing of the Khaltan anticlinal, in the section along the Khata-dere river. In such cases in the transitive packet the red colour of rocks may often be observed. The Nardaran series is represented in the flysh facies by an alternation of greenish-grey calcareous clays, marls, whitish limestones and calcareous sandstones. Occasionally lensiform streaks of conglomerates are observed. The thickness of the series in the Kizil-Kazma section amounts to 350 m, that along Isti-su-dere—650 m.

In the Besh-Barmak subzone the lower part of the Nardaran series is saturated with block conglomerates, containing large inclusions of Lusitanian limestones and shales with *Posidonia buchi* Roem. In marl calcareous packets are present aptychi and ammonites of the group *Spiticeras negreli* Math., *Berriasella calisto* d'Orb., *Perisphinctes* (*Aulacosphinctes*) *transiformis* Opp. (accurate determination was impossible owing to poor preservation of the fauna).

According to N. P. Luppoff, to whom belongs the determination of the fauna, the age of the strata should be referred to Tithonian—Lower Valanginian.

It is known that almost allover Caucasus the transition from Tithonian to Valanginian proceeds gradually. One does, therefore, not always succeed in discriminating both layers in case characteristic fauna is absent. On the other hand, we know that Upper Tithonian transgression is regionally spread in a number of districts.

As with the Nardaran series begins a new transgressive complex of deposits, then, according to what was said above, the age of the lower horizons of this series can be safely reduced to the boundary between Upper and Lower Tithonian.

Kaytar series. In full conformity, Tithonian-Valanginian series of rocks (Nardaran) is overlain by a new series, possessing a more terrigene character and formed chiefly of dark clays with layers of other rocks. Of these the brownish-yellow ferruginous marls and tutensteins are the most typical. Besides that, occur lenses of spherosiderites.

In the Kaytar series, having some resemblance with the lower siderite beds of the north-eastern Caucasus, Hauterivian belemnites are found (ex gr. *Duvalia bipartita* Blainv.).

The upper boundary of the series is determined quite arbitrarily by the occurrence of more light and more carbonatic clays and by the disappearance of resemblance with the lower siderite beds of the n.-w. Caucasus.

The thickness of the Kaytar series (Hauterivian) is about 250 m.

Khalchai series. Powerful series of grey and olive-grey marlaceous clays with separate horizons, containing large inclusions, sometimes whole cliffs of malm limestones, is connected by a gradual transition with under and overlying deposits. By far less important are the thin layers of calcareous alevrolites, occasionally bearing traces of submarine syngenetic deformation, with hieroglyphs and with a structure of the Fontainebleau type. In upper parts of the series occurs small fauna of ammonites, brachiopods, gastropods etc. According to Mishunina, the fauna is of transitional character from Lower Aptian to Barremian. The upper most horizon of the Khalchai series contains secretions of limestones with veins of calcite. In Soviet-abad in the upper parts of the series occur numerous *Orbitolina*.

The thickness of the series is about 900—1100 m. The Khalchai series is overlain conformably by the Khanagja series (Aptian) and Alty-agach (Lower and Mid-Albian).

The comparison of our scheme of the Lower and Upper Jurassic stratigraphy with those by K. M. Bogdanovich and more recent schemes by M. F.Mirchink, V. E. Khain, I. M. Goubkin and S. F. Fedorov is given in the table.

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GENETICS

ON THE CHROMOMERE STRUCTURE OF WHEAT CHROMOSOMES

By N. T. KAKHIDZE

(Communicated by N. I. Vavilov, Member of the Academy, 4.IX. 1938)

The discoveries made with respect to the structure of the chromosomes in cells of the salivary glands of *Drosophila* show beyond doubt that behind the morphology of chromosomes, as usually seen, there is hidden a still more detailed structure. The finest structural elements known to us in plant chromosomes are the chromomeres. At meiosis they are regarded as the most elementary morphological units of which the chromosomes are composed ([1],[2]). Although a considerable number of papers have treated the subject of the chromomere structure of chromosomes, only very few investigators have succeeded in individualizing the chromomeres so as to be able to give a morphological characteristic of the meiotic chromosomes [([3]), in part([4]) and([5])].

Fig. 1.

The present investigation had as its aim to approach the study of the chromomere structure of wheat chromosomes at meiosis. A form of durum wheat, *Triticum durum* var. *hordeiforme*, was taken for investigation. From material fixed with a 5 : 5 chrom-formalin solution ([6]), having first dipped the anthers in alcohol, there were obtained quite a number of anthers, in which at pachynema it was possible to distinguish the chromomeres composing the chromosomes. Microtome sections were cut 10 microns thick and stained with gentian-violet. The use of microtome sections has the disadvantage that the long, twisted pachytene threads are cut in two, and it is impossible to trace them over a great distance. Hence, methods should be devised for bringing out the chromomeres in whole nuclei—by smears, which has not yet been possible to achieve with wheat.

In order to avoid any difficulty as regards the identification of the chromosomes studied, those chromosomes were chosen which had a specific relationship to the nucleolus. At the pachynema stage two of the bivalents are always attached to the nucleolus (Fig. 6, 8). At the present time there is no doubt that the process of accumulation of nucleolar substance is linked with a definite region in these chromosomes—the secondary constriction region, and that, consequently, this region is in contact with the nucleolus([7]).

These secondary constrictions, thus functionally differing from other secondary constrictions, are, moreover, in contrast to the latter, readily brought out in mitotic chromosomes also by the use of Navashin's fixative(6).

According to McClintock(8), in *Zea Mays* directly adjoining the satellite thread—both at mitosis and at meiosis—there is a deep-staining chromatin body, which, in her opinion, is active in the process of the formation of the nucleolus. At metaphase among the chromosomes of durum wheat there are two particular pairs with distinct secondary constrictions (unpublished data from the Cytological Laboratory of the Institute of Plant Industry), and at mitotic prophase there are found four chromosomes attached to the

Fig. 2.

Fig. 3.

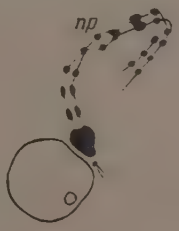

Fig. 4.

Fig. 5.

nucleolus just as described in the paper of Heitz(7), Kaufmann(9) and McClintock(8): the secondary constriction passes through the entire nucleolus and the appendage appears on the opposite side of the nucleolus (Fig. 1); in the figure the chromosomes are separated one from another). In *Zea Mays* the union of the chromosomes with the nucleolus at meiosis is in prin-

Fig. 6.

Fig. 7.

Fig. 8.

ciple the same as at mitosis, with the difference, however, that the secondary constriction does not pass through the whole nucleolus, the bivalents being attached to the nucleolus only by the «nucleolar-organizing body». The reason for this difference is unclear and requires special study.

In *Triticum durum* studied by us, as in *Zea Mays*, the bivalents are attached to the nucleolus by means of a hetero-chromatin body, from which it may be concluded that the bivalents in *Triticum* are also attached to the nucleolus at a point directly adjoining the secondary constriction. The hetero-chromatin body attached to the nucleolus is double, and at the pachynema stage it looks like two large chromomeres («contact chromomeres») (Fig. 2—8). From these data it may be concluded, first, that the two bivalents of *T. durum* attached to the nucleolus are those four chromosomes which at the mitotic prophase have appendages, and, secondly, that the place of attachment of the bivalents to the nucleolus is constant. Hence, when subjecting to analysis the sections adjoining the place of attachment to the nucleolus, we may be

assured that we are studying the same sections. The contact chromomeres of the two pairs of bivalents attached to the nucleolus markedly differ in size (Fig. 6, 8). These loci are also used as starting points, when counting and noting the characteristics of the chromomeres in these chromosome sections.

In the nucleolus, opposite to each pair of contact chromomeres, are peculiar «vacuoles», differing from one another in size according to the difference in size of the contact chromomeres. By careful adjustment of the microsope these «vacuoles» are clearly visible and appear to be something like canals leading from the chromomeres into the nucleolus. It is possible that these «vacuoles» have some connexion with the metabolism of the nucleolus, and this should be taken into consideration in a special study of the latter question. It may besides be presumed that these «canals» are secondary constrictions, embedded in the nucleolus, which would account for the absence of any connexion on the surface of the latter. In our case the «vacuoles» are used only to fix the loci, at which the bivalents are attached to the nucleolus.

We shall take the size of the contact chromomeres as the first distinguishing character in the structure of the bivalents. The bivalent with the larger contact chromomeres we shall denote by the letter *A*, and the one with the smaller contact chromomeres by the letter *B*. In our study of the chromomeres we analyzed sections on either side of the contact chromomeres, these sections constituting only a small part of the entire chromosome thread. For convenience in our preliminary description we shall arbitrarily call the section on one side the «right» and the one on the other the «left» one.

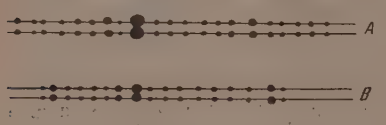

Fig. 9—10.

In Fig. 2, 3, 4 and 5 are shown the two sections of bivalent *A*, while in Fig. 9 is given a schematic drawing of the order of their chromomeres. All the figures are at pachynema of cells from anthers of a single plant. The structure of these is as follows. In the «right» section lie first (proceeding from the contact chromomeres) five pairs of small chromomeres, then one pair of somewhat larger ones, then a pair of considerably larger, and, finally, several pairs of small chromomeres (Fig. 3, 4, 5). In the «left» section, immediately adjoining the contact chromomeres, is a pair of very small chromomeres (Fig. 2 and 5), which in some cases we are unable to detect (Fig. 3); then comes a pair of large chromomeres; then five pairs of small ones (Fig. 2), of which the first two sometimes appear somewhat larger; and next a pair of medium-sized chromomeres. Beyond these chromomeres the chromosome thread continues, if at all, but for a very short distance, the given section apparently being the distal end of the chromosome, i. e. the appendage separated from the main body of the chromosome by the secondary constriction.

The other bivalent *B*, attached to the nucleolus, is characterized by the following structure. The contact chromomeres, although also large, are considerably smaller than those of bivalent *A*. In this case we did not succeed to individualize the chromomeres either. In the «right» section we were able to distinguish nine pairs of chromomeres, of which seven were of about the same size, while the eighth pair was much larger and the ninth small again (Fig. 8). In the «left» section next to the contact chromomeres is a pair of quite large chromomeres, then three small pairs, a still smaller one, one large pair, and one medium-sized pair (Fig. 6 and 7).

Taking into account the importance of the problem as to the genesis of species of wheat, the analyses of chromosome structure may also be uti-

lized to this purpose. Since in the somatic set there are two pairs of chromosomes with appendages, even in the case of their not being entirely alike morphologically, the question arises as to the degree of resemblance of their structure to judge whether these chromosomes point to a double set (autotetraploid). A comparison of the sections of *A* and *B* adjoining the secondary constriction (Fig. 9 and 10) shows a difference in the chromomere structure of these sections in the two bivalents. Whether or not this difference may be accounted for by the non-homology of the chromosomes or a regrouping of the chromomeres (from inversions, etc.), can be made clear only after an analysis of the chromomere structure of the entire chromosomes is performed.

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CYTOLOGY

THE EFFECT OF CONVALLARINE UPON THE SEEDS OF SUMMER WHEAT

By A. S. AFANASSIEVA

(Communicated by N. I. Vavilov, Member of the Academy, 21.IX.1938)

The investigations of Lits showed in 1934 that colchicine arrests mitosis at its beginning. Brues(2), while studying the effect of colchicine in order to find out whether its action is stimulating or inhibitory, has demonstrated the absence of any stimulating effect of colchicine. The two investigators have studied the influence of colchicine on the animal tissues. The tissue of the liver served as a material. In 1937 there appeared the work by Blakeslee and Avery(1) on the influence of colchicine upon plants; it raised a great interest among geneticists and plant-breeders. The interest could be justified by the fact that Blakeslee and Avery succeeded in obtaining by the application of colchicine abundant polyploid forms, tissues and isolated cells in various plants; this is not only of theoretical interest, but of practical importance too. D. Kostoff(3) has also subjected to the influence of colchicine and acenaphthene seeds of various plants (*Triticum*, *Secale*, *Zea*, *Cannabis*, and others). The occurrence of polyploid cells (tetra-, octoploid and of higher grades of polyploidy) under the influence of the two substances was strikingly apparent; increase of the number of nuclei that were sometimes of unequal size and often deformed like those in cancer tumors was observed, the number and sometimes the size of the nucleoli was increased. M. Navashin(4) in studying the influence of acenaphthene upon the division of the nucleus and cell in a number of plants has observed in *Crepis capillaris* the formation of polyploid cells as well as various irregularities in cell division, including multiprotoplastic cells which are being repeatedly observed by me in wheat after the treatment with X-rays.

It was of interest to try the effect of other substances on the plant cell. Following the advice of professor S. L. Ivanov, I have chosen convallarine.

Seeds of summer wheat *Triticum vulgare* var. *caesium* 0111 served as a material for the investigation. The following solutions of convallarine were prepared for the treatment of seeds: 0.1, 0.2, 0.5, 0.8 and 1.0%. The actual convallarine content in the solution was lower that that indicated above, since the solubility of convallarine in water is not complete. A precipitate is deposited from the solution, the latter is turbid and forms foam when shaken. The seeds were placed in small Petri dishes, 25 per dish, on filter paper, and soaked in corresponding solutions of convallarine.

Soaking lasted for four days. Seeds soaked in distilled water served as control, the duration of soaking being the same. After soaking the condition of seeds kept at different concentrations proved different. The majority (70—80%) of the control seeds as well as of those subjected to the action of 0.1, 0.2 and 0.5% solutions started germinating; in some of the seeds sub-

jected to the influence of 0.8 and 1.0% solutions the coat has begun to burst slightly, while the remainder failed to germinate by that time. After soaking all the seeds were placed into quartz sand, in order to obtain rootlets for cytological investigations.

The ratio of plants produced proved also different depending on the concentration of the solution. Thus, seeds soaked in 0.1, 0.2 and 0.5% solutions gave 80—88% of plants, while only 60% of seeds treated with 0.8 and 1.0% solutions of convallarine produced any plants at all. Rootlets from all the plants were fixed, Navashin's fixative having been applied.

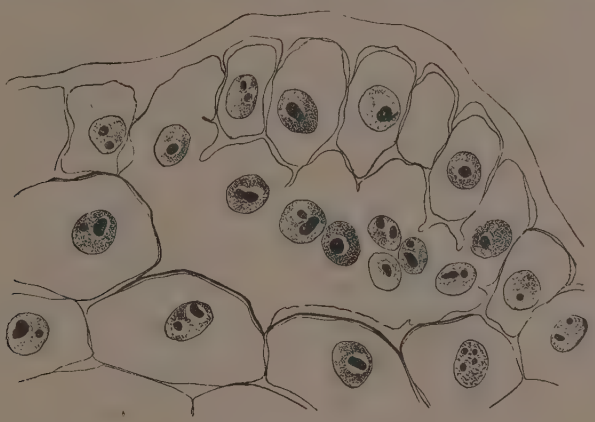

Fig. 1. The syncytium.

Sections 14 microns thick were made and stained with iron haematoxylin after Heidenhain.

Cytological investigations yilded the following result. No alterations were observed in cells of rootlets which had developed from seeds soaked in 0.1, 0.2 and 1.0% solutions of convallarine; there was noted but a large amount of mitoses. Only two concentrations of convallarine (0.5 and 0.8% solutions) had produced certain alterations in the cell. To such alterations belongs the formation of a syncytium which is a partial result of cytomyxis. One of the instances of formation of a syncytium is represented on Fig. 1. [All the drawings (Fig. 1, 2, 3) were made with the aid of an immersion objective 90× and an eyepiece 7×].

The figure represents that part of the cross-section of the root, where the syncytium has the appearance of a kind of neoplasm or a tumour located between the epidermis and the second cell layer of the periblem. The walls of the periblem cells which adjoin this neoplasm are highly thikened as if fenced off from it. The syncytium contains a large accumulation of nuclei immersed into a common mass of cytoplasm where all the cell walls have disappeared, and only here and there at the boundary of the epidermis remnants of cell walls can be observed. This syncytium is apparent even under low power and has the appearance of a small protuberance. The very beginning of the formation of such a syncytium failed to be observed, although in another plant we succeeded in observing its appearance, while tracing it under the microscope from section to section. In one of the sections of the rootlet a binucleated cell is distinguished through its size. This cell becomes still larger in the next section. A single nucleus is located in its centre, a kind of rudiments of partition walls are observed at both sides of it. The nucleus of the one of the cells of epidermis is pulled toward this large cell and is

emitting a part of its nuclear substance into the cytoplasm, being ready for creeping over into this cell. This beginning of cytomyxis may be observed in Fig. 2. The aspect is changed in the next section. A similar large cell of the same shape with an scarcely noticeable partition wall in the middle is observed here, but the number of nuclei has increased to five.

It may be suggested that a division of the nucleus without cellular division takes place in the cell; the possibility of the nuclei creeping over from the neighbouring cells is not excluded either; the number of nuclei in the cell is increased by both processes. The syncytium represented in Fig. 3 is

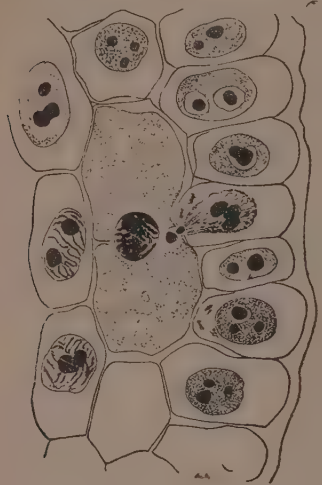

Fig. 2. The beginnng of cytomyxis.

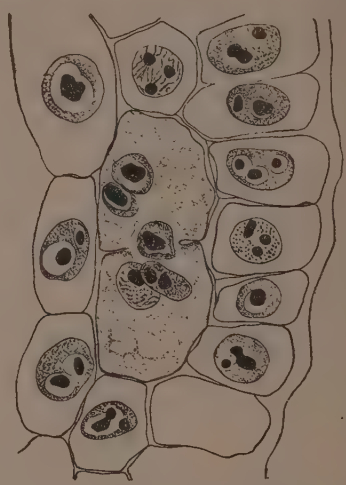

Fig. 3. The beginning of the origination of a syncytium.

thus formed. In the following sections of this giant cell the fusion of nuclei resulting in a decrease in their number is observed, thereupon their number is increased again. Toward the end of the slide, i. e. in the more differentiated part of the meristem of the root, the nidus of the syncytium formation disappears and two wholly normal cells are observed instead. While the nidus of the syncytium is extinguished, it is formed again in another part of the rootlet. The phenomenon of cytomyxis and the formation of a syncytium was observed only in four plants, two of which developed from seeds soaked in 0.5% solution of convallarine, and the remaining two—from seeds soaked in 0.8% solution. In the first instance this corresponds to 11.7% of plants with changed cells, in the second—to 20.0%.

It may be thus seen that out of the five concentrations of convallarine tried, only two cause alterations in the cell. These alterations bear deleterious character and may lead to the formation of tumours.

The investigation on the effect of convallarine shows that the alkaloid exerts an unquestionable influence upon the cell; it produces, however, no polyploid cells and cannot be, therefore, applied for the production of polyploid plants with the same success as colchicine and acenaphthene.

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HYDROLOGIE

EINE QUANTITATIVE DREDGE ZUR ERFORSCHUNG DER BENTHONISCHEN BESIEDELUNG DER PFLANZENBESTÄNDE EINES WASSERBECKENS

Von W. J. BUT

(Vorgelegt von A. A. Richter, Mitglied der Akademie, d. 22.IX.1938)

Während den Fragen der quantitativen limnischen Untersuchung der Bodenwelt des Wasserbeckens viel Interesse entgegengebracht wird, bleibt die Vorschungsmethode für die benthonische Besiedelung seiner Pflanzenbestände fast ganz aus, was im wesentlichen durch das Fehlen irgendwelcher Apparatur zur quantitativen Auswertung dieser Region zu erklären ist.

Die bis jetzt bekannten Versuche die Tierwelt der Pflanzenbestände in Süss- und Meerwasserbecken zu erforschen wurden mit Hilfe einer recht primitiven Zählmethode ausgeführt.

Korotun(2) gebrauchte bei der Untersuchung der Tierwelt des Pflanzenwuchses einen Käscher von 30 cm Durchmesser, indem er mit 2 Schwungbewegungen 4 m lange Wasserbahn durchmachte und die Individuenzahl pro 0.25 m² bestimmte. Somow(5) empfiehlt die Makrophytenzone auf einer Oberfläche von 1² Arsch. mit Hilfe eines Käschers zu untersuchen und mit gleichmässigen Bewegungen die Tiere des Bodens und des Pflanzenwuchses abzufangen. Wodjanitsky (4) untersuchte den Pflanzenwuchs in 1 m Seetiefe; er benutzte eine Flasche (15 cm Durchmesser, 35 cm hoch) ohne Boden, bedeckte damit die *Cystosira*-Büsche und schnitt dann dieselben von unten ab. Segerstrale(6) untersuchte die Tierwelt in *Fucus*-Beständen: er sammelte die Vegetation in einen Käscher und rechnete dann die erbeuteten Tiere auf 1 kg *Fucus* um.

Während meiner Untersuchung der Biozönosen und des Zoobenthos der Pflanzenbestände in den Wasserbecken des Überschwemmungsgebietes, habe ich in 1936 eine Dredge zur quantitativen Auswertung der benthonischen Organismen der Pflanzenzone konstruiert.

Das Konstruktionsprinzip der Dredge ist das Herausschneiden eines bestimmten Wasservolumens mit allen inbegriffenen Pflanzen und Tieren; es ist also demselben der Methode der Zählquadrate nahe, zwar mit Berücksichtigung der dritten Dimension. Das Herausschneiden einer bestimmten Wassersäule wird mit Hilfe der sich oberhalb und unterhalb eines Zylinders bewegenden Messer erreicht. Diese Wassersäule stellt nach ihrer Form und ihrem Volumen einen Viertel des Zylinders dar.

1. Beschreibung des Apparates. Wie es aus der Abbildung und dem Photo zu sehen ist, ist der Körper der Dredge von zwei flachen und einer gebogenen Wand begrenzt; die letztere stellt einen Viertel der Oberfläche eines Zylinders vor.

Der schneidende Mechanismus besteht aus einem unteren und einem oberen beweglichen Messer und zweier Verbindungsplatten, die diese Messer an einem Ende festhalten. Das andere Ende dieser Messer bewegt sich frei um eine Achse. Die Verbindungsplatten sind von beiden Seiten der gebogenen Wand im 1—2 mm Abstand von derselben angebracht. An der inneren Platte ist ein Rand des Dredgenbeutels befestigt, die äussere trägt zwei Ohren für die zum Schliessen der Dredge dienende Schnur und einen Handgriff zur Handhabung der Messer.

Die erste flache Wand hat von aussen ein mit einer Klemmschraube versehenes Rohr zum Festhalten der Stange, und zwei grosse und zwei kleine Rollen mit Scheiben, die die Leine zurechthalten. Innerhalb dieser Wand sind an ihrem unteren und oberen Rand die unbeweglichen Messer befestigt. Innerhalb der zweiten flachen Wand ist an ihren drei Wänden der Dredgenbeutel angenäht und zwei Schafte zum Aufsetzen der Messer angebracht. Der Dredgenbeutel ist aus Kongresstoff hergestellt und hat unten eine schliess-

Fig. 1. Eine quantitative Dredge für Pflanzengewuchs (halbgeschlossen).

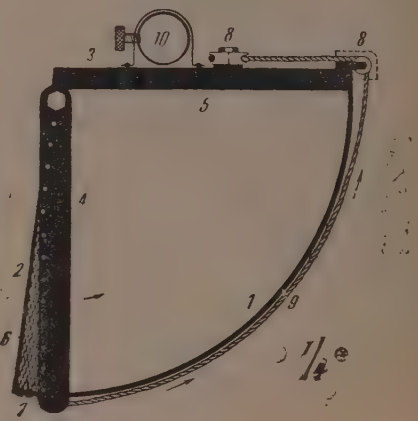

Fig. 2.

bare Öffnung. Längs der gebogenen Wand ist er durch zwei flache bewegliche, die Wand umschliessende Ringstücke unterstützt. Bei der offenen Dredge liegt der Beutel innen, neben der zweiten flachen Wand zusammengefaltet und von den beweglichen Messern vor Verletzungen geschützt. Die Stange ist hölzern und je 25 cm markiert. Diese Zeichen gestatten das Einstellen der Dredge in beliebiger Schicht des Pflanzenbestandes (in gewünschter Höhe von der Oberfläche oder vom Boden an). Der Verschluss der Dredge erfolgt durch eine von den beweglichen Messern ausgehende Schnur, die mit Hilfe einiger Rollen längs der Stange gerichtet wird und an ihrem oberen Ende befestigt ist. Schematisch ist die Dredge in der Abb. 2 dargestellt (offen und von oben gesehen, Volumen 1/80 m³). *1*—die gebogene Wand (innere Länge—480.7 mm); *2*—die zweite flache Wand (234.7 mm);*3*—die erste flache Wand (235.0 mm); *4*—das obere bewegliche Messer mit Löchern zum Annähen des Dredgenbeutels (30 mm breit); *5*—das obere unbewegliche Messer (20 mm breit); *6*—der zusammengefaltete Beutel; *7*—die Beutelstützerflache Ringstücke; *8*—Rollen mit Scheiben (15 und 35 mm Durchmesser); *9*—die Schnur; *10*—das Rohr mit einer Klemmschraube (40 mm Durchmesser) für die Stange.

Die Masszahlen für die Anfertigung der Dredge wurden nach folgenden Formeln berechnet: R (Dredgenradius)=235.0 mm (vom Drehungspunkt der Messer bis zur inneren Seite der gebogenen Wand gemessen); f (Kreisfläche)=

$\pi R^2 = 2206.18$ cm^2: f_1 ($^1/_4$ Kreisfläche) $= \frac{2206.18}{4} = 551.55$ cm^2; f_2 ($^1/_2$ Fläche des beweglichen Messers) $= 15$ mm x 265 mm $= 38.25$ cm^2; f_3 (Fläche des Ineinandergreifens des beweglichen Messers und des unbeweglichen) $= 13.30$ cm^2; F (Fläche des Fangraumes der Dredge) $= f_1 - f_2 - f_3 = 551.55$ cm$^2 - 38.25$ cm$^2 - 13.30$ cm$^2 = 500.0$ cm$^2 = {}^1/_{20}$ m^2; H (Dredgenhöhe) $= 25.0$ cm; V (Dredgenvolumen) $= F \cdot H = 500.0$ x $25.0 = 12\,500$ cm$^3 = {}^1/_{80}$ m^3. Das Gewicht der Dredge mit 3 m langer Stange beträgt 7000 g.

2. Gebrauchsweise. Um eine Probe in der Pflanzenwuchszone zu nehmen wird die Dredge in Offenzustand vertikal ins Wasser versenkt und zwar in der Weise, dass sie die unterhalb der Dredge befindlichen Pflanzen umkleidet. In der gewünschten Tiefe wird die Dredge mit Hilfe der an der Stange befestigten Schnur zugemacht, damit wird ein bestimmtes Volumen aus einer bestimmten Vegetationsschicht ausgeschnitten und das erbeutete im Dredgenbeutel eingeschlossen. Darauf wird die Dredge herausgezogen, die Schnur unter dem Beutel losgebunden und der Fang in eine emaillierte Cuvette von 24 x 30 cm Inhalt gebracht. Dann spülen wir die Dredge mit einem auf die Aussenwand des Beutels gerichteten Wasserstrahl aus und legen den ganzen Inhalt der Cuvette in ein 1 Liter fassendes Gefäss.

Die erste Untersuchung des Materials wird an Ort und Stelle gemacht, die Tierorganismen (das Benthos der Pflanzenwuchszone) werden in Gefässen von 200—300 cm^3 Inhalt aufgehoben. Die Pflanzen müssen entweder sofort nach den Arten geordnet und abgewogen oder ins Laboratorium gebracht und dort in trockenem Zustande auf ihr Trockengewicht geprüft werden.

3. Charakteristik des Apparates. Die beschriebene Dredge eignet sich um von einem für 3—4 Menschen bestimmten Boote aus bedient zu werden. Während der Untersuchung des Sees Snytkowo im Herbst 1936 mit Hilfe dieser Dredge von $^1/_{80}$ m^3 Volumen und einer 3 m langen Stange, konnte die ganze Arbeit von einem Menschen durhgeführt werden.

Die Schnittleistung der Dredge wurde an verschiedenen Pflanzen geprüft; es erwies sich, dass nur 7—8 cm dicke *Typha*-Stengel und dichter, alter *Phragmites*-Bestand schwer zu schneiden waren, übrigens arbeitete das Gerät unfehlbar.

Bei der Entnahme der Proben aus der Zone halbuntergetauchter Vegetation ist anzuraten zunächst die oberen Teile der Stengel oberhalb des Wassers mit einer Scheere abzuschneiden und erst dann die Dredge zu benutzen.

Zum Schluss möchte ich manche Angaben anführen, die für die Arbeit mit der Dredge charakteristisch sind. Mit Hilfe dieser Dredge von $^1/_{80}$ m^3 Volumen wurde in 1936 im See Snytkowo (im Überschwemmungsgebiet, in der Umgebung der Hydrobiologischen Donetz-Station Zmiew) das Benthos der Makrophytenbestände untersucht. Aus einer nach der Zusammensetzung und Dichtigkeit der Besiedelung einheitlichen Assoziation wurden aus derselben Schicht 10 gleiche Proben entnommen; von solchen Assoziationen wurden 10 mit verschiedener Zusammensetzung untersucht. Die Ergebnisse dieser Untersuchung werden in einem besonderen Artikel dargelegt (3), an dieser Stelle führen wir nur einige Tatsachen an, die unsere Dredge als einen bei der Untersuchung von Biozenösen des Makrophytenbestandes gut verwendbaren Apparat kennzeichnen. Das Fangvermögen des Apparates kommt durch Erbeuten so flinker Schwimmer wie *Plea minutissima* F. zum Vorschein. *Plea* hat in allen Assoziationen des Wasserbeckens 22% Frequenz, in den Biozönosen *Lemnaetum* und *Typhaetum* ist sie sogar 70—100%. *Ilybius subaeneus* Er. hat im *Potamogetonetum* und *Sparganietum* 20—30% Frequenz. Die Larven von *Agrion puella* L. haben die allgemeine Frequenz von 25%, in einzelnen Assoziationen aber 20—70%. *Argyroneta*

aquatica L. zeigt in einzelnen Assoziationen bis 40% Frequenz; die Insektenlarven—30—40%.

Was die Anzahl einheitlicher Proben bei den Orientierungsuntersuchungen anbelangt, so sehen wir aus den Tabellen der quantitativen und qualitativen Beschaffenheit der Assoziationen(3), dass wir bei vier einheitlichen Proben aus derselben Vegetationsschicht im Verglelch mit 10 Proben nur geringe Abweichungen in den Arten- und Quantitätswerten bemerken. Die Anzahl einheitlicher Proben, die zur Charakteristik eines Benthos oder einer Schicht in der Biozönose erforderlich sind, kann uns die Tabelle 1 zeigen, wo die Frequenz der Tiergruppen in 10 verschiedenen Assoziationen des Sees Snytkowo angegeben ist. In dieser Tabelle ist die Frequenz folgenderweise berechnet: bei 1 Probe ist sie im Verhältnis zu 100 Proben aus der Makrophytenzone bestimmt; bei 2 Proben wurden die Werte von benachbarten Proben der Artentabelle der Biozönose summiert und als eine einzige betrachtet, es ergaben sich 50 solche Proben; bei 3 Proben wurde die Berechnung auf dieselbe Weise, d. h. im Verhältnis zu 33 Proben, gemacht; bei 4 Proben—im Verhältnis zu 25 Proben. Der mittlere Frequenzfehler wird nach der Formel $\frac{\sqrt{\%P_0 - \%P_1}}{\sqrt{N}}$ bestimmt, wo $\%\ P_1$—die Frequenz, $\%\ P_0$—Proben ohne Organismen, N—die Anzahl der Proben [(1), S. 300] bezeichnen.

Tabelle 1

Frequenz P_1 der Tierkomponente in 10 verschiedenen Biozönosen der Makrophytenzone (Pflanzenwuchs-Benthos) eines Sees (Snytkowo) des Überschwemmungsgebietes

	1 Probe, %	Doppelte Probe, %	Dreifache Probe, %	Vierfache Probe, %
Oligochaeta	26±4.4	50±7.0	67±8.2	80±8.0
Hirudinea	50±5.0	82±5.4	87±5.9	92±5.4
Mollusca	95±2.2	100±0.0	100±0.0	100±0.0
Malacostraca	45±5.0	64±6.8	74±7.7	88±6.5
Acarina	46±5.0	84±5.2	90±5.3	96±3.4
Odonata	52±5.0	70±6.5	80±7.0	92±5.4
Ephemeroptera	96±2.0	100±0.0	100±0.0	100±0.0
Trichoptera	81±3.9	94±3.3	97±3.0	100±0.0
Rhynchota	27±4.4	38±6.8	56±9.7	65±9.5
Coleoptera	36±4.8	58±6.8	86±6.1	92±5.4
Chironomidae (larvae)	90±3.0	100±0.0	100±0.0	100±0.0

Aus dieser Tabelle ist ersichtlich, dass falls in einer zur Charakteristik der Gruppen entnommenen Probe die Frequenzzahlen der meisten Tiergruppen unter 100% stehen, so betragen sie bei 4 Proben 100% oder sind diesem Werte nahe.

Die Frequenz einzelner Arten bei 1 und bei 4 Proben ist in der Tabelle 2 dargestellt.

Diese Tabelle ist ebenso berechnet, wie die Tabelle 1, d. h. die Frequenz ist im Verhältnis zu allen Biozönosen berechnet, während viele Arten an einzelne Biozönosen angewiesen sind und in anderen gar nicht vorkommen. Wenn wir zum Beispiel die Frequenz bei 1 und bei 4 Proben in einzelnen Biozönosen beobachten, so ergiebt sich für die meisten Arten 90—100% Frequenz (im Vergleich mit 10 Proben). So hat *Leptocerus aterrimus* St. im *Sparganietum* und auf submersen Blättern im *Sagittarietum* 100% Frequenz (bei 4 Proben); *Plea minutissima* F.—im *Lemnetum* und *Typhaetum*—100%; *Planorbis vortex* L.—im *Lemnetum* 90%, im *Ceratophylletum* und *Myriophylletum* — 100% u. s. w. Daraus ist zu folgern, dass die Ergebnisse

Tabelle 2

	1 Probe, %	4 Proben, %
Stylaria lacustris L.	26±4.4	65±4.7
Herpobdella octoculata L.	43±4.9	83±3.7
Planorbis vortex L.	62±4.8	92±2.7
Planorbis vorticulus Tr.	23±4.2	49±5.0
Limnaea ovata Dr.	54±5.0	83±3.7
Physa fontinalis L.	44±5.0	79±4.1
Bithynia tentaculata L.	42±4.9	72±4.4
Bithynia leachi Shepp.	37±4.8	74±4.4
Ancylus lacustris L.	40±4.9	58±4.9
Viviparus viviparus L.	36±4.8	60±4.9
Asellus aquaticus L.	42±4.9	68±4.7
Cloëon dipterum L.	95±2.2	100±0.0
Caenis dimidiata St.	48±5.0	67±4.7
Plea minutissima F.	22±4.1	40±4.9
Setodes tineiformis Curt.	59±4.9	89±3.1
Triaenodes bicolor Curt.	36±4.8	47±4.2
Leptocerus aterrimus St.	21±4.1	29±4.5
Ilybius (larvae)	22±4.1	50±5.0
Gliptotendipes sp. *Gripekoveni*	71±4.5	100±0.0
Endochironomus sp. *Signatornis*	38±4.8	70±4.6
Endochironomus sp. *Nymphoides*	33±4.7	61±4.9

bei 4 einheitlichen Proben aus einer Assoziation nur sehr wenig von solchen bei 10 Proben abweichend sind, dabei werden die zufälligen Arten natürlich nur zu 40—50% berechnet.

Schlüsse

1. Die beschriebene Dredge gibt uns die Möglichkeit die quantitativen Proben aus verschiedenen Horizonten des Makrophytenbestandes zu entnehmen und seine Besiedelung, Produktion, Beschaffenheit der Biozönosen u. s. w. zu untersuchen.

2. Für die allgemeine Charakteristik des Pflanzenwuchsbenthos müssen wir nicht weniger als 4 einheitliche Proben aus jedem Punkt entnehmen. Die Anzahl der Proben, die bei der Entscheidung einzelner Fragen der Biozönologie der Makrophytenzone nötig sind, ist teilweise aus den in der Arbeit von But (3) angeführten Tabelle ersichtlich.

3. Wenn an der unteren Öffnung der Dredge ein Metallnetz eingesetzt wird, so kann die Dredge zur Probeentnahme in der Zone der schwimmenden Pflanzen benutzt werden. Der Apparat wird unter der Vegetation eingestellt und im Offenzustand hochgezogen.

4. Beim Einsetzen des Netzes an der oberen Öffnung der Dredge kann die ganze Vegetationssäule mit ihrer Tierwelt auf der $^1/_{20}$ m² Oberfläche entnommen werden.

5. Bei der Herstellung des Apparates aus Duraluminium kann das Volumen und die Fangfläche, sowie die Genauigkeit der Wirkungsweise der Dredge in manchen Biozönosen erhöht werden.

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